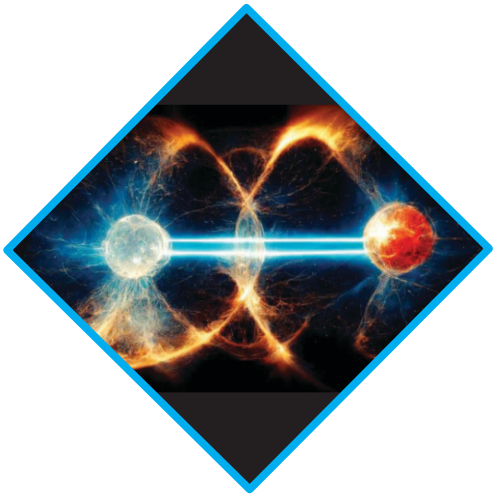
TEST EDITION

THE TEXTBOOK OF

PHYSICS

For Class XII



Sindh Textbook Board

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PREFACE

The century we have stepped in, is the century of Science and technology. The modern disciplines of Physics are strongly influencing not only all the branches of science but each and every aspect of human life.

To keep the students abreast with the recent knowledge; it is must that the curricula at all the levels be updated. Moreover regularly by introducing the rapid and multidirectional development taking place in all the branches of Physics.

The recent book of Physics for Class - XII has been written in this preview and in accordance with the revised curriculum. Prepared by Ministry of Education, Govt of Sindh. Reviewed by independent team of Directorate of Curriculum Assessment and Research, Jamshoro Sindh. Keeping in view the importance of Physics, the topics have been revised and re-written according to the need of the time.

Among the new editions the introductory paragraphs, information boxes, summaries and a variety of extensive exercises have been included. Which I think will not only develop the interest but also add a lot to the utility of the book.

The Sind Textbook Board has taken great pains and incurred expenditure in publishing this book inspite to its limitations. A textbook is indeed not the last word and there is always room for improvement. While the authors have tried their level best to make the most suitable presentation, both in terms of concept and treatment. There may still have some deficiencies and omissions. Learned teachers and worthy students are therefore requested to be kind enough to point out the short comings of the text or diagrams and to communicate their suggestions and objections for the improvement of the next edition of this book.

In the end, I am thankful to our Authors, Editors and Subject specialist of Board for their relentless service rendered for the cause of education.

Chairman
Sindh Textbook Board

Unit 15

Molecular Theory of Gases

Teaching Periods 07 **Weightage %** 05



Popcorn is a fun way to learn about the kinetic molecular theory of gases, the phase change of water from a liquid to a gas. When heated, the water inside turns to steam, making the kernel pop and expand up to 50 times its size.

In this unit student should be able to:

- Recall concept of temperature
- Solve problems using scales of temperature and their conversion
- Explain triple point of water
- State general gas law
- Derive gas laws (Boyle's law, Charle's law and Avogadro's law)
- Solve problems using gas laws
- State the basic postulates of KTG
- Describe the molecular movement causes the pressure exerted by gas and derive pressure equation
- Describe the relation between kinetic energy of molecules and temperature
- Solve problems using relation between kinetic energy and temperature

Introduction:

We have based our study of Mechanics on the three fundamental quantities mass, length and time. Now we need to introduce the additional fundamental quantity temperature. We have used physical quantities such as total mass m, number of moles n, pressure P, Volume V and temperature T to describe the quality and condition of the material. We now want to pull several relationships together into more general formulation of the behavior of materials. A central concept in this synthesis is equation of state of a material. We begin with a general discussion of the molecular structure of matter then we develop the kinetic Molecular Model of an ideal gas, deriving the pressure of an ideal gas on the basis of kinetic Molecular Model.

15.1.1 Temperature:

Temperature is a measure of the average translational kinetic energy of the molecules of body.

How does Thermometer device help to measure the temperature of any other body? The answer to this question leads to the concept of thermal equilibrium. When two bodies at different temperatures are brought in thermal contact with each other, the heat start flowing from the hot body to the cold body till the temperature of both bodies becomes same then they are said to be in thermal equilibrium. The S.I Unit of temperature is Kelvin.

15.1.2 Scales of Temperature:

There are three scales of temperature

(i) Centigrade or Celsius scale
(ii) Fahrenheit scale
(iii) Kelvin or absolute scale are commonly used these days

DO YOU KNOW?

A temperature scale that starts at absolute zero, the theoretical lowest possible temperature. It is measured in Kelvin (K),

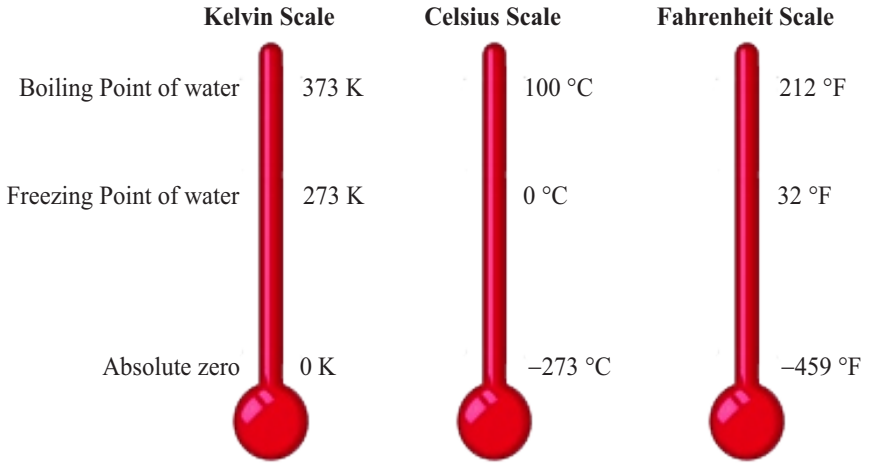


Figure 15.1
Scales of temperature mentioning the different temperatures.

Centigrade or Celsius Scale (Anders Celsius):

In the Celsius scale the freezing point of water or Melting point of ice is marked 0^0C and boiling point is 100^0C. The interval between these two points is divided into hundred equal parts. Each part thus represents one degree Celsius (1°C).

(i) Fahrenheit Scale (Daniel Fahrenheit):

In Fahrenheit scale the fixed points are marked 32°F and 212°F respectively and the interval between the fixed points is divided into 180 equal parts. Each part represents 1°F.

(ii) Kelvin Scale (James Lord Kelvin):

In this scale the melting point of ice is 273K and the boiling point of water is 373K. The temperature is given in units called Kelvin instead of degrees. The lowest temperature is 0K known as absolute zero.

In order to derive empirical formulae among Centigrade, Fahrenheit and Kelvin scales, let the three thermometers be placed in a bath tub and the mercury in each thermometer rises to the same level as shown in figure 15.1. We arrive at the relation.

$$\frac{\text{Temp: on one scale} - \text{F.P}}{\text{B.P} - \text{F.P}} = \frac{\text{Temp: on 2nd scale} - \text{F.P}}{\text{B.P} - \text{F.P}}$$

$$\frac{T_C - 0}{100} = \frac{T_F - 32}{180} = \frac{T_K - 273}{100} \quad \ldots\ldots\ldots\ldots (15.1)$$

Centigrade temperature can be converted into Kelvin temperature by simply adding 273 to the centigrade temperature i.e.

$$T_K = 273 + T_C$$

DO YOU KNOW?

(i) The scale divisions on the Kelvin scale are equal in size to the divisions on Celsius scale. i.e 1k K= 1°C.
(ii) 1°C = 1.8°F

DO YOU KNOW?

If two systems are each in thermal equilibrium (equal temperature) with a third system, then the two initial systems are in thermal equilibrium with each other.

DO YOU KNOW?

(i) Fahrenheit and Celsius scales coincide at -40°
(ii) Fahrenheit and Kelvin scales coincide at 574.25 or 574.25K

Worked Example 15.1

The normal human body temperature is 36.88°C. What is this temperature on Fahrenheit scale?

Solution:

Step 1: Write the known quantities and point out quantities to be found

T_C = 36.88°C

T_F?

Step 2: Write the empirical formula and rearrange if necessary

$$\frac{T_C}{5} = \frac{T_F - 32}{9}$$

$$T_F = \frac{9}{5} T_C + 32$$

Step 3: **Put the values and calculate.**

$$T_F = \frac{9}{5} \times 36.88 + 32$$

$$T_F = 98.4°F$$

Result: $T_F = 98.4°F$

Self-Assessment Questions:

1. Convert each of the following temperature from the centigrade scale to Kelvin scale and Fahrenheit 0°C, 20°C, 120°C, 500°C, –23°C, 200°C.
2. Convert each of the following temperature from the Kelvin scale to the Celsius scale and Fahrenheit 0K, 20K, 100K, 300K, 373K, and 500K.

15.1.3 The Triple Point of Water:

The vapor, liquid, and solid states of water can coexist in thermal equilibrium only at a specific pressure and temperature.

The triple point of water occurs at a specific temperature and pressure, where all three phases are in thermodynamic equilibrium. This means that at the triple point, ice, liquid water, and water vapor can exist together without any phase changing to another.

These conditions are used to define the Kelvin temperature scale, where the temperature at the triple point of water is precisely 273.16 K (0.01°C) and 4.58 mm of mercury (611.73 Pa) and is used to calibrate thermometer.

In a figure 15.3, the triple point is represented as the point where the lines separating the solid, liquid, and gas phases meet. This diagram visually demonstrates the relationship between temperature and pressure for the different phases of water.

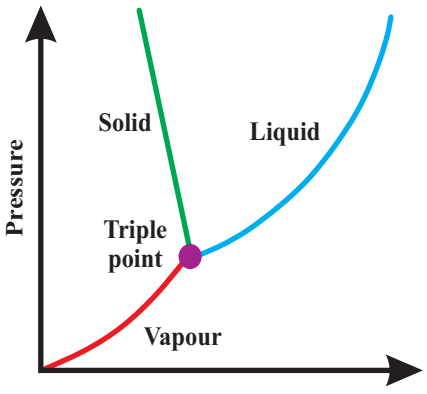


Figure 15.3 A triple point of water

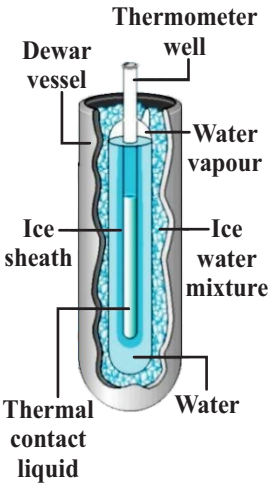


Figure 15.2
A triple point cell in which solid ice liquid water are water vapour co-exist in thermal equilibrium. The bulb of the constant volume gas thermometer is shown inserted into the well of the cell.

The temperature at which Triple point of water has been set by international agreement to be T_3= 273.16K. In T_3 the subscript 3 means "Triple Point" this agreement also sets the size of the Kelvin as $\frac{1}{273.16}$ of the difference between the Triple point temperature of water and absolute zero.

Self-Assessment Questions:

1. What are the conditions (temperature and pressure) at which the triple point of water occurs?
2. Where is the triple point located in relation to the solid, liquid, and gas regions?

15.2 Gas Laws

Gases have no fixed volume or shape and their volume at same mass can be altered by changing the pressure as well as the temperature. Gases can be described with the help of four variables i.e. pressure, volume, mass and temperature. The relation between any two variables is found experimentally while keeping the other two constant.

15.2.1: Boyle's Law

Boyle's law states that "Volume V of given mass of a gas is inversely proportional to the pressure P, provided the temperature T of the gas remains constant."

Boyle's law can be written as:

$$V \propto \frac{1}{P} \text{ (at constant temperature)}$$

$$PV = \text{constant} \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots \quad (15.2)$$

We can also represent Boyle's law on a graph, as shown in Fig: 15.4(a) The graph plotted between P and V at constant temperature is a curve called hyperbola showing the inverse relation between them for two different states, while fig. 15.4(b) graph of P plotted against $\frac{1}{V}$ is a straight line passing through the origin, showing direct proportionality. Boyle's law can be written as:

$$P_1V_1 = P_2V_2 = \text{Constant} \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots \quad (15.3)$$

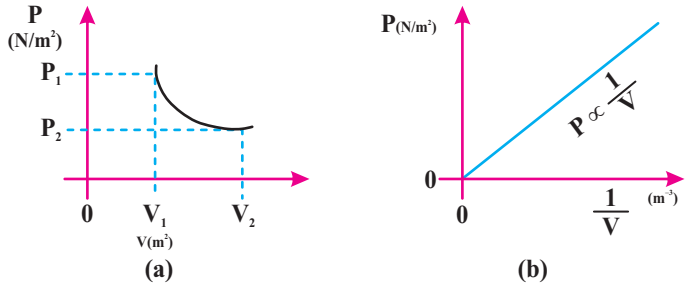


Figure 15.4 (a) and (b)
showing the relation between pressure and volume respectively

DO YOU KNOW?

"Liquid vapor equilibrium region" corresponds to the end point at the top of vaporization curve in fig: 15.4, it is called critical point and the corresponding values of P and T are called critical pressure Pc, and temperature Tc. A gas above the critical temperature does not separate into phases when it is compressed isothermally.

Charles's Law:

This law states that "The volume V of a given mass of a gas is directly proportional to the absolute temperature T at constant pressure P".

Charles law can be written as

$V \propto T$ (at constant pressure)

Or $\frac{V}{T}$ = constant …… (15.4)

If V_1 and V_2 are the volumes of the gas at temperature T_1 and T_2 respectively then for two different states, Charles's law is represented as

$$\frac{V_1}{T_1} = \frac{V_2}{T_2} = \text{constant} \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots (15.5).$$

From the graph (fig: 15.5) at 0°C the gas still has a volume V_0. The graph between volume and temperature is a straight line. If the graph is extrapolated backward, it cuts the temperature axis at –237°C.This graph shows that volume of a gas is zero. Kevin selected this temperature (–273°C) as the zero called absolute zero (0K).

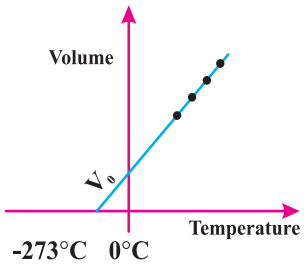


Figure 15.5
Shows relation between temperature and volume

DO YOU KNOW?

The temperature at which volume of a gas becomes zero and molecular motion ceases is termed as absolute zero.

Avogadro's Law:

In 1811, Italian scientist Amedeo Avogadro suggested his hypothesis regarding the relationship between volume and number of molecules of a gas. This hypothesis now called Avogadro's law which states that "equal volume of all gases contains the same number of molecules at the same temperature and pressure". Thus, the volume of a gas is directly proportional to number of moles of the gas at constant temperature and pressure.

In symbol, we can write as:

$V \propto n$ (at constant temperature and pressure)

$$\frac{V}{n} = \text{constant} \ldots\ldots\ldots\ldots (15.6)$$

Where V is volume and n is the number of moles of a gas. Thus $1dm^3$ (or cm^3, m^3) of oxygen contains the same number of molecules as $1dm^3$ or $1\ cm^3$) etc of hydrogen or of any other gas, provided the volumes are measured under the same conditions of temperature and pressure.

It can be written as

$$\frac{V_1}{n_1} = \frac{V_2}{n_2} = \text{constant} \ldots\ldots\ldots\ldots\ldots\ldots\ldots (15.7).$$

When V_1 and V_2 are volumes of gas and n_1 and n_2 are amount of gas.

DO YOU KNOW?

Avogadro's number:
A mole of any substance is that mass of substance that contains a specific No. of molecules called Avogadro's number.

$N_A = 6.022 \times 10^{23} \frac{\text{molecules}}{\text{mole}}$

Avogadro's number is defined to be the No. of carbon atoms in 12g of the isotope carbon-12. The number of moles of a substance is related to its mass m. $n = \frac{m}{M}$

Where M is the molecular mass of the substance usually expressed in g/mole.

Example:
Blowing up a balloon is an example of Avogadro's law, because it says as you blow more molecules of air into the balloon it expands.

15.2.2 General Gas Law:

In order to derive general gas law, we make use of Boyle's law, Charles's law and Avogadro's law. An interrelation among the physical quantities e.g. pressure, volume, temperature and amount of matter of a given sample of gas is termed as "equation of state" of gas or General gas law.

According to Boyle's law:

$V \propto \frac{1}{P}$ (when n number of mole and temperature T are kept constant)

According to Charles's law:

$V \propto T$ (when n and pressure P are kept constant)

According to Avogadro's law:

$V \propto n$ (when T and P are kept constant)

Consider for a moment that none of the variable are to be kept constant, then all the above three relationships can be joined together.

$$V \propto \frac{nT}{P}$$

$$V = constant \times \frac{nT}{P}$$

$$V = \frac{R\,n\,T}{P} \quad \ldots\ldots\ldots\ldots\ldots\ldots \quad (15.8)$$

Where **R** is constant of proportionally and is called General gas constant or universal gas constant and does not depend on the quantity of gas in the sample. If P is measured in Nm^{-2} V in m^3 and T in Kelvin then the volume of universal gas constant is **R = 8.314J mol^{-1}, k^{-1}**

Above equation is written as:

$PV = nRT$ …… (15.9)

DO YOU KNOW?

Real or permanent gas is a gas that obeys gas Laws at high temperature and low Pressure only.

DO YOU KNOW?

1dm = 0.1m

Hence $1dm^3 = 10^{-3}m^3$

Ideal or perfect gas is a gas that obeys gas Laws at all temperatures and pressures.

Worked Example 15.2

A cylinder contains 0.80 dm^3 of nitrogen gas at a pressure of 1.2 atmosphere (1 atm = 1.01 10^5 N/m^2). A piston slowly starts compressing the gas to a pressure of 6.0 atm. The temperature of the gas remains constant. Calculate the final volume of the gas. (The temperature of the gas is constant and that its mass is fixed).

Solution:

Step 1: Write the known quantities and point out quantities to be found.

$P_1 = 1.2$ atm, $V_1 = 0.80 dm^3$, $P_2 = 6.0$ atm, $V_2 = ?$

We don't worry about the particular units of pressure and volume being used here, so long as they are the same on both sides of the equation. The final value of V_2 will be in dm^3, because V_1 is in dm^3

Step 2: Write down the formula and rearrange if necessary.

$P_1V_1 = P_2V_2$

Substitute the values in equation and calculate.

$1.2 \text{ atm} \times 0.8 dm^3 = 6.0 \text{ atm} \times V_2$

$$V_2 = \frac{1.2 \text{ atm} \times 0.8 dm^3}{6.0 \text{ atm}} = 0.16 \text{ } dm^3$$

Result: The volume of the gas is reduced to 0.16dm3or 1/5th of original volume so the pressure increases by a factor of 5.

Worked Example 15.3

Find the mass of the air within the car tyre, given that it holds 0.020m³ of air at 27°C and a pressure of 3.0×10^5 N/m². The molecular mass of air is 28.8g/mol.

Solution:

Step 1: Write the known quantities and quantities to be found.

$P = 3.0 \times 10^5$ N/m²
$T = 27°C = 300K$
$V = 0.020m^3$
Molar mass of air = 28.8g/mol.
R = 8.314J mol–1 k–1
i. No. of moles of air = n =?
ii. Mass of the air = m =?

Step 2: Write down the formula and rearrange if necessary.

$$n = \frac{PV}{RT}, \quad = \frac{3.0 \times 10^5 \times 0.020}{8.314 \times 300} = 2.41 \text{mol}$$

Step 3: Now we calculate the mass of air.

mass = number of moles × molar mass
m = 2.41 mol × 28.8g/mol = 69.4g ≅ 69g

Result: i. 2.41 mol ii. 69g

Self-Assessment Questions:

1. How does the pressure of a gas change if its volume is halved at constant temperature?
2. What happens to the volume of a gas when its temperature is increased while keeping the pressure constant?
3. What is the relationship between the volume and the number of moles of a gas at constant temperature and pressure?

15.3 Kinetic theory of gases:

We discussed the properties of an ideal gas by using such macroscopic variables such as pressure, volume and temperature. Such large-scale properties can be related to a description on a microscopic scale, where matter is treated as collection of molecules. Now we begin by developing a microscopic model of an ideal gas called Kinetic Theory of gases. The first step in the construction of a theory is to setup some sort of model which is simple enough to be treated analytically.

15.3.1 Basic Postulates of Kinetic Theory of Gases:

The basic postulates of Kinetic Theory of gases are as under

1. A gas contains a very large number of particles called molecules. Depending on the gases each molecule consists of an atom or a group of atoms.
2. A finite volume of a gas consists of very large number of molecules. This assumption is justified by experiments. At standard conditions there are 3×10^{25} molecules per cubic meter.
3. The size of the molecules is much smaller than separation between molecules; it is about 3×10^{-10}m.

4. The molecules move in all directions with various speeds collide elastically with one another and with the walls of the container.
5. The molecules exert no forces on one another except during collisions. In the absence of the external forces, they move freely in straight lines.
6. Laws of mechanics are assumed to be applicable to the motion of molecules.

15.3.2 Pressure of Gas:

The pressure exerted by a gas is merely the momentum transferred to the walls of the container per second per unit area due to the continuous collisions of molecules of the gas. In order to calculate the pressure of an ideal gas from Kinetic Theory. Let us consider a cube having side length L whose walls are perfectly elastic contains N number of molecules each of mass m as shown in figure 15.6.

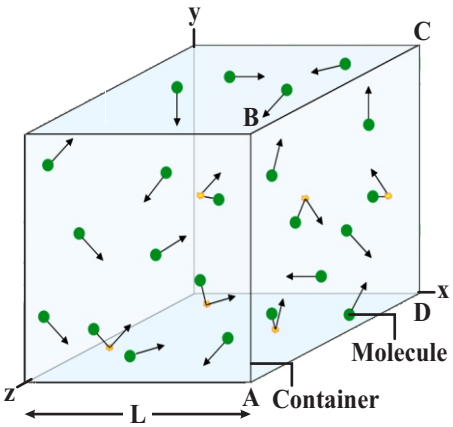


Figure 15.6
Shows gas molecules moving in random directions within container.

Consider a single molecule of mass m moving with velocity V_1parallel to x-axis.It moves back and forth, colliding at regular intervals with the ends of the box and thereby contributing to the pressure of the gas. A molecule which has a velocity V_1 can be resolved into three rectangular components V_1x, V_1y and V_1z parallel to three co-ordinates axis x, y and z.

A molecule which collides with the face ABCDA of the cube, it will rebound elastically in opposite direction, such that x-component of the velocity V_1x, is reversed, the V_1yand V_1z remain unaffected. Therefore, the momentum before collision is mV_1x and after collision is -m V_1x causing a change of momentum.

Change in momentum= P_i-P_f = =mV_1x – (- mV_1x) =mV_1x + mV_1x

Change in momentum = $2mV_1x$ …… 15.10

After recoil the molecule travels to opposite face and collides with it, rebounds and travels back to the face ABCDA after covering a distance 2L. The time Δt between two successive collisions with face ABCDA is:

$$\Delta t = \frac{2L}{V_1x} \quad \ldots\ldots\ldots\

Similarly, the forces due to all other molecules can be determined. Thus, the total x-directed F due to N number of molecules of the gas moving with velocities V_1, V_2, V_3,……,V_n is:

$$F = F_1 + F_2 + F_3 + \ldots\ldots\ldots\ldots + F_n$$

$$F = \frac{mV_1^2\varkappa}{L} + \frac{mV_2^2\varkappa}{L} + \frac{mV_3^2\varkappa}{L} + \ldots\ldots\ldots\ldots + \frac{mV_n^2\varkappa}{L}$$

As pressure is normal force per unit area, hence pressure P on the face perpendicular to x-axis is:

$$P = \frac{F}{A} = \frac{F}{L^2}$$

$$P = \left(\frac{mV_1^2\varkappa}{L} + \frac{mV_2^2\varkappa}{L} + \frac{mV_3^2\varkappa}{L} + \ldots\ldots\ldots + \frac{mV_n^2\varkappa}{L}\right)/L^2$$

$$P = \frac{m}{L^3}(V_1^2\varkappa + V_2^2\varkappa + V_3^2\varkappa + \ldots\ldots\ldots + V_n^2\varkappa) \ldots\ldots\ldots (15.12)$$

The number of molecules in unit volume n_v is $\frac{N}{L^3}$. Where N is the total number of molecules. Therefore:

$L^3 = \frac{N}{n_v}$ and substituting this value in eq: (15.12)

$$P = mn_v\left(\frac{V_1^2\varkappa + V_2^2\varkappa + V_3^2\varkappa + \ldots\ldots\ldots + V_n^2\varkappa}{N}\right) \ldots\ldots\ldots (15.13)$$

Where mn_v is the mass per unit volume which we call density ρ and $\frac{V_1^2\varkappa + V_2^2\varkappa + V_3^2\varkappa + \ldots\ldots\ldots + V_n^2\varkappa}{N}$ is the average value of $\overline{V_x^2}$. for all the molecules in the container, we call this average square velocity $\overline{V_x^2}$. The square root of $\overline{V_x^2}$. is referred as V_{rms}. Eq: 15.13 can be written as:

$$P = \rho\overline{V_x^2} \ldots\ldots\ldots\ldots\ldots (15.14)$$

The terms $\overline{V_x^2}$. is only one component of the total velocity. Since $\overline{V^2} = \overline{V_x^2} + \overline{V_y^2} + \overline{V_z^2}$ on the average. $\overline{V_x^2} = \overline{V_y^2} = \overline{V_z^2}$ due to randomness of the molecular motion.

$\overline{V^2} = 3\overline{V_x^2}$. and $\overline{V_x^2} = \frac{1}{3}\overline{V^2}$

Substituting this value into the above equation, we find that:

$$P = \frac{1}{3}\rho\overline{V^2} \ldots\ldots\ldots\ldots\ldots (15.15)$$

15.3.3 The relation between Kinetic Energy of Molecules and Temperature:

The relation between molecular Kinetic Energy and temperature can be derived by using eq: 15.15

$$P = \frac{1}{3}\rho\overline{V^2}$$

$$P = \frac{1}{3}mn_v\overline{V^2} \text{ since } \ell = (mn_v) \ldots\ldots\ldots\ldots (15.16)$$

Since n_v represents the number of molecules per unit volume $n_v = \frac{N}{V}$ Eq: (15.16) can be written as

$$PV = \frac{1}{3}m\frac{N}{V}\overline{V^2} \ldots\ldots\ldots\ldots (15.17)$$

Now we can compare the equation (15.17) with this ideal gas equation PV = nRT. The left-hand sides are the same. So, the two right-hand sides must also be equal.

$$\frac{1}{3}Nm\overline{V^2} = nRT \quad \ldots\ldots\ldots\ldots (15.18)$$

Substituting $n = \frac{N}{N_A}$ and multiplying both sides by $\frac{3}{2}$, we obtain the relation.

$$\frac{3}{2}\times\frac{1}{3}Nm\overline{V^2} = \frac{3}{2}\times\frac{N}{N_A}RT$$

$$\frac{1}{2}m\overline{V^2} = \left(\frac{3}{2}\right)\left(\frac{R}{N_A}\right)T$$

Since $\frac{R}{N_A} = k$ (Boltzmann constant)

$k = 1.38 \times 10^{-23}$ J/Molecule k

Hence $\frac{1}{2}m\overline{V^2} = \frac{3}{2}kT \quad \ldots\ldots\ldots\ldots (15.19)$

$K.E_{avg} = \frac{3}{2}kT$

The mean translational Kinetic energy of a molecule of an ideal gas is proportional to the absolute temperature.

Hence, mean translational Kinetic energy of a molecule $\propto$ T.

DO YOU KNOW?

K (Boltzmann constant) is universal gas constant per molecule of a gas R is a universal gas constant per mole of a gas.

Worked Example 15.4

The atoms in a gas have mean translational (K.E) equal to 5.0×10^{-21}J. Calculate the temperature of the gas in K and °C.

Solution:

Step 1: Write the known quantities and point out quantities to be found

Mean K.E = 5.0×10^{-21}J

K = 1.38×10^{-23}J/molecule. K

(i). T_k = ?

(ii). T_c = ?

Step 2: Write down the formula and rearrange if necessary.

$$\text{Mean K.E} = \frac{3}{2}kT$$

$$T = \frac{\text{mean K.E}}{k}\times\frac{2}{3} = \frac{5.0\times10^{-21}}{1.38\times10^{-23}}\times\frac{2}{3}$$

T = 242K

Thus, temperature in centigrade scale is:

T = K – 273

T = 242 – 273 = – 31°C.

Result: **(i).** 242K, **(ii).** –31°C.

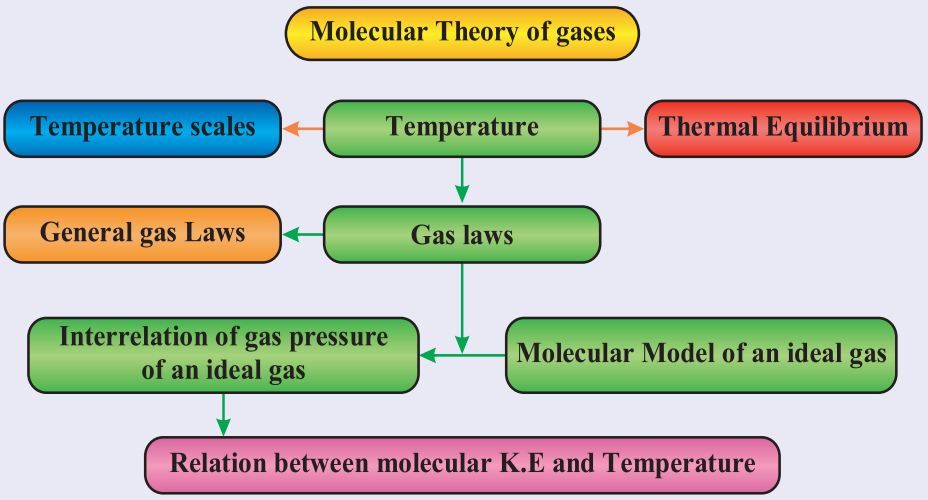
Self-Assessment Questions:

1. Describe the motion of molecules in a gas according to the kinetic molecular theory.
2. What is the relationship between temperature and molecular motion in a gas?



- ✓ Temperature is a measure of the average kinetic energy of particles in a substance.
- ✓ The triple point of water is the temperature (0.01°C) and pressure (611.73 Pascal's) at which water coexists in all three phases (solid, liquid, and gas).
- ✓ The general gas law states that PV = nRT, relating pressure, volume, number of moles, and temperature of a gas.
- ✓ Boyle's Law: $P_1V_1 = P_2V_2$ (pressure and volume are inversely proportional at constant temperature)
- ✓ $V_1/T_1 = V_2/T_2$ (volume and temperature are directly proportional at constant pressure)
- ✓ V = RT/P (volume and number of moles are directly proportional at constant temperature and pressure)
- ✓ KTG (Kinetic Theory of Gases) Postulates:
  1. Gases consist of tiny particles (molecules) that are in constant motion.
  2. The molecules are very small compared to the distance between them.
  3. The molecules are in constant random motion.
  4. The molecules collide with each other and the container walls.
- ✓ Molecular movement and pressure: The pressure exerted by a gas is caused by the collisions of molecules with the container walls, and is related to the temperature and volume of the gas by the equation P = (2/3) nkT.
- ✓ The kinetic energy of molecules is directly proportional to the temperature of the gas, and is described by the equation KE = (3/2) kT.

CONCEPT MAP



EXERCISE

Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. The relationship between temperature and average kinetic energy of particles in a gas is:
   (a) temperature is inversely proportional to average kinetic energy
   (b) temperature is directly proportional to average kinetic energy
   (c) temperature is independent of average kinetic energy
   (d) temperature is proportional to the square of average kinetic energy

2. Standard conditions of temperature and pressure (STP) refer to a gas at:
   (a) 0°C and 1 atm (b) 20°C and 1 atm
   (c) 25°C and 1 atm (d) 30°C and 101.3 kPa (1 atm)

3. If the temperature is kept constant and the volume of a gas is doubled, then pressure of a gas is:
   (a) Reduced to $\frac{1}{2}$ of the original value. (b) Doubled
   (c) Reduced to $\frac{1}{4}$ of the original value (d) Quadrupled

4. The Avogadro's number is the number of molecules in:
   (a) One mole of a substance (b) One kg of a substance
   (c) One m^3 of a gas (d) One kilogram of hydrogen gas

5. Mean translational K.E. per molecule of an ideal at temperature T is:
   (a) $\frac{3}{2}kT$ (b) $\frac{1}{2}kT$ (c) $\frac{2}{3}kT$ (d) kT^4

6. The normal human body temperature is:
   (a) 98.6°F (37°C) (b) 99.6°F (37.4°C)
   (c) 100.4°F (38°C) (d) 101°F (38.3°C)

7. The pressure P, the density ρ and the average speed of molecules of an ideal gas are related by the equation.
   (a) $P = \frac{2}{3} m\ell^2$ (b) $P = \frac{1}{3} m\ell^2$ (c) $P = \frac{1}{3}\rho\overline{V^2}$ (d) $P = \frac{2}{3}\ell^2$

8. In air at S.T.P, the average speed of the:
   (a) Oxygen molecules is greater than Nitrogen molecules
   (b) Nitrogen molecules is greater than Oxygen molecules
   (c) Oxygen molecules is approximately equal to Nitrogen molecules
   (d) Helium atoms is greater than both Oxygen and Nitrogen molecules

9. If the absolute temperature of a gas is increased 3 times, the rms velocity of the molecules will be:
   (a) 3 times (b) 9 times (c) $\sqrt{3}$ times (d) $\frac{1}{3}$ times

10. A gas is enclosed in an isolated container which is placed on a fast-moving train uniformly. The temperature of the gas:
    (a) Increases due to the motion of the train
    (b) Decreases due to the motion of the train
    (c) Remains constant
    (d) Fluctuates, depending on the train's speed and direction

Section (B): CRQs (Short Answered Questions):

15.1. Why the earth is not in thermal equilibrium with the sun?
15.2. Describe the relationship between temperature and kinetic energy of molecules.
15.3. It is observed that when mercury in glass thermometer is put in a flame, the column of mercury first descends and then rises. Explain.
15.4. What is standard temperature, pressure?
15.5. A thermometer is placed in direct sun light. What will it read the temperature?
15.6. The pressure in a gas cylinder containing hydrogen will leak more quickly than if it is containing oxygen. Why?
15.7. When a sealed thermos bottle full of coffee is shaken, what are the changes occur?

15.9. How does the Kinetic theory account for the following observed facts:
(a) A gas exerts pressure
(b) The pressure of a gas depends upon its temperature.

15.10 Calculate the average speed of an air molecule at room temperature (20°C) and compare it to the speed of sound in air (330 m/s).

Section (C): ERQs (Long Answered Questions):

15.1. What is temperature? Explain the scales of temperature in detail.
15.2. Define and explain Boyle's law, Charles's and Avogadro's law.
15.3. Derive general gas law by making use of gas laws.
15.4. Describe the molecular movement causes the pressure exerted by gas, derive pressure equation.
15.5. Interpret mathematically that temperature is a measure of average translational K.E of the molecules of a gas.

Section (D): Numerical:

15.1. The freezing point of mercury is –39°C.Convert it into °F and the comfort level temperature of 20° into Kelvin. **(Ans: –38.2°F, 293K)**

15.2. The boiling point of liquid nitrogen is -321°F. Change it into equivalent Kelvin temperature. **(Ans: 77K)**

15.3. Calculate the volume occupied by a gram-mole of a gas at 0°C and a pressure of 1.0 atmosphere. **(Ans: 22.4 liters/mole)**

15.4. An air storage tank whose volume is 112 liters contain 3kg of air at a pressure of 18 atmospheres. How much air would have to be forced into the tank to increase the pressure to 21 atmospheres, assuming no change in temperature? **(Ans: 0.5kg)**

15.5. A balloon contains $0.04m^3$ of air at a pressure of 120KPa. Calculate the pressure required to reduce its volume to 0.025 m^3 at constant temperature. **(Ans: 1.9×10^5Pa)**

15.6. The molar mass of nitrogen gas N_2 is $28gmol^{-1}$. For 100g of nitrogen, calculate.
(a) the number of moles. **(Ans: (a) 3.57 mole)**
(b) the volume occupied at room temperature (20°C) and pressure of 1.01×10^5 Pa. **(Ans: (b) $0.086m^3$ or $86dm^3$)**

15.7. A sample of a gas contains 3.0×10^{24} atoms. Calculate the volume of the gas at a temperature of 300K and a pressure of 120K Pa. **(Ans: $0.104m^3$)**

15.8. Calculate the root mean square speed of hydrogen molecules at 0°C and 1.0 atm pressure. Assuming hydrogen to be an ideal gas. The density of hydrogen is 8.99×10^{-2} kg/m^3. **(Ans: $1835.86ms^{-1}$)**

15.9. Calculate the root mean square speed of hydrogen molecule at 500K (mass of proton = 1.67×10^{-27}kg and K = 1.38×10^{-23}J/molecule.-K) **(Ans: $2489.49ms^{-1}$)**

15.10. (a) Determine the average value of the Kinetic energy of the particles of an ideal gas at 10°C and at 40°C. **(Ans: $5.86x10^{-21}$J, $6.48x10^{-21}$J)**
(b) What is the Kinetic energy per mole of an ideal gas at thesetemperatures? **(Ans: 3526.57J, 3901J)**

Unit 16

First Law of Thermodynamics

Teaching Periods 08 **Weightage %** 05



The Thar Coal Power Plant, located in the Thar Desert of Sindh, Pakistan, is a significant project aimed at utilizing the large reserves of lignite coal found in the region. The Thar Coal Power Plant is an example of how the first law of thermodynamics is applied in real-world energy systems.

In this unit student should be able to:

- Describe that heat flow and work are two forms of energy transfer between systems and calculate heat being transferred.
- Relate rise in temperature of a system increases its internal energy.
- Explain that internal energy is determined by the state of the system and that it can be expressed as the sum of the random distribution of kinetic and potential energies associated with the molecules of the system.
- Define thermodynamics and various terms associated with it.
- Calculate work done by a thermodynamic system during a volume change.
- Describe the first law of thermodynamics expressed in terms of the change in internal energy, the heating of the system and work done on the system.
- Explain that first law of thermodynamics expresses the conservation of energy.
- Describe the applications of first law of thermodynamics with the help of equations and graphs
- Solve the problems using the equations of first law of thermodynamics
- Define the terms specific heat and molar specific heats of a gas.
- Apply first law of thermodynamics to derive Cp – Cv =R.
- Solve the problems using equations of specific heat

Introduction:

In this unit, we explore key concepts in thermodynamics: heat, work, internal energy, the First Law of Thermodynamics, and molar specific heat. These concepts form the basis for understanding energy transfer and conservation within systems. Heat and work are fundamental modes of energy transfer, influencing changes in a system's internal energy. This unit provides a comprehensive overview, essential for grasping energy transformations and their applications in diverse fields.

16.1.1 Heat and Work:

Work and thermal energy are linked by the broader framework of energy conservation. The study of heat has evolved significantly over centuries. As shown in table 16.1.

Table 16.1: History of heat energy

18th Century	Isaac Newton	➢ Proposed the concept of "caloric" (a hypothetical fluid thought to be the substance of heat). ➢ Influenced future research.
18th Century	Joseph Black	➢ Introduced the concept of latent heat. ➢ Distinguished between heat and temperature.
18th Century	Count Rumford (Benjamin Thompson)	➢ Challenged the caloric theory of heat ➢ Demonstrated that heat is a form of motion, not a material substance
19th Century	Sadi Carnot	➢ Laid the foundations of thermodynamics ➢ Described the Carnot cycle
19th Century	James Prescott Joul	➢ Established the mechanical equivalent of heat ➢ Showed that energy can be converted from one form to another
19th Century	Rudolf Clausius and William Thomson (Lord Kelvin)	➢ Developed the concept of entropy ➢ Formulated the second law of thermodynamics
20th Century	Albert Einstein	➢ Provided evidence for the atomic theory of matter ➢ Helped establish the kinetic theory of heat
Modern Developments:		➢ Advances in statistical mechanics and quantum mechanics ➢ Connected heat to the behavior of atoms and molecules

These discoveries laid the foundation for modern thermodynamics, which explains heat as the kinetic energy of particles in motion.

Work can lead to changes in the internal energy of a system, contributing to thermal energy. The first law of thermodynamics encapsulates this relationship, emphasizing the interplay between heat transfer, work and changes in internal energy within a system.

Thermal Energy and Work:

Internal Energy and Heat (Q): In thermodynamics, thermal energy is associated with the internal energy of a system. Heat (Q) is the transfer of thermal energy between systems due to a temperature difference.

Heat Flow:

It flows spontaneously from a region of higher temperature to one of lower temperature. The rate at which heat is transferred can be quantified using the equation:

$$Q = mc\Delta T \ldots\ldots\ldots 16.1$$

Where Q is the heat transferred, m is the mass of the substance, c is the specific heat capacity of the substance, and ΔT is the temperature change.

Above equation, known as the heat transfer equation, highlights the dependence of heat transfer on the material properties and the extent of temperature change.

Conversion of Work to Thermal Energy:

Work done on a system can lead to an increase in its internal energy, contributing to thermal energy. For instance, when you rub your hands together, the mechanical work done generates heat due to friction, increasing the thermal energy of your hands.

The concept of PV work can be illustrated by imagining a gas in a cylinder equipped with a movable piston as shown in figure 16.1. The pressure exerted on the piston is essentially a force per unit area. When the gas undergoes expansion, the piston moves upward, necessitating energy input (a force acting over a distance). The work done can be expressed as the product of the force and distance. In the case of a gas in a cylinder (thermal system), this simplifies to the pressure multiplied by the change in volume.

$$W = P\Delta V \ldots\ldots\ldots 16.2$$

$$W = -P\Delta V$$

The negative sign is to keep our sign convention for energy going into or out of the system.

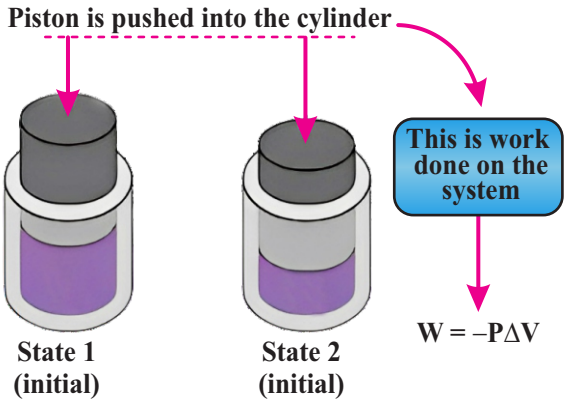


Figure 16.1 Work done by PV relation.

Self-Assessment Questions:

1. **Define heat is a form of energy.**
2. **Explain the basic concept of how work can be converted into thermal energy.**

16.2.1 Internal Energy:

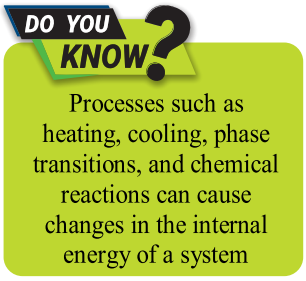
Energy can generally be classified into two forms: kinetic and potential energy. The molecules of all substances contain both kinetic and potential energies. The amount of kinetic and potential energy of a substance depends on the phases of matter (solid, liquid or gas), this is known as the **internal energy**

The internal energy of a substance is defined as:

"***The sum of the random distribution of kinetic and potential energies within a system of molecules.***"

The symbol for internal energy is *U*and its SI unit is **Joule** (J).

- The internal energy of a system is determined by:
  - Temperature
  - The random motion of molecules
  - The phase of matter: gases have the highest internal energy, solids have the lowest
- The internal energy of a system can increase by:
  - Doing work on it
  - Adding heat to it
- The internal energy of a system can decrease by:
  - Losing heat to its surroundings



Internal Energy & Temperature:

The internal energy of an object is intrinsically related to its temperature. When a gas in a container is heated, the gas molecules move faster, increasing their kinetic energy. In a solid, where molecules are tightly packed, heating causes the molecules to vibrate more. In both liquids and solids, molecules have both kinetic and potential energy due to intermolecular forces that keep them close together. However, ideal gas molecules are assumed to have no intermolecular forces, meaning they only possess kinetic energy and no potential energy.The (change in) internal energy of an ideal gas is equal to:

$$\Delta U = \frac{3}{2} k\Delta T \ldots\ldots\ldots 16.3$$

Therefore, the change in internal energy is proportional to the change in temperature:

$$\Delta U \propto \Delta T$$

Where:ΔU = change in internal energy (J)
ΔT = change in temperature (K)

"As the container is heated up, the gas molecules move faster with higher kinetic energy and therefore higher internal energy". as shown in the figure 16.2.

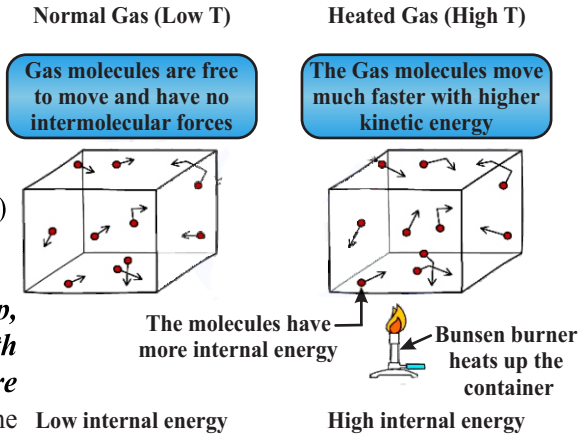


Figure 16.2 Internal Energy & Temperature

Self-Assessment Questions:

1. How the internal energy of an object is related to its system?
2. How does an increase in temperature affect the internal energy of a substance?

16.3.1: Define thermodynamics and various terms associated with it.

Thermodynamic terms and basic concepts:

The study of the flow of heat or any other form of energy into or out of a system as it undergoes a physical or chemical transformation is called thermodynamics.

A thermodynamic system is a specific portion of the physical universe that is chosen for analysis. It is defined by boundaries and the surrounding. ***The boundary is the physical or imaginary separation between the system and its surroundings. The surroundings refer to everything outside the boundaries of the system.***

This includes the rest of the universe that can interact with the system. separates the system from the rest of the universe. The system can be of any size or shape and can consist of a single substance or a combination of substances as shown in figure 16.3.

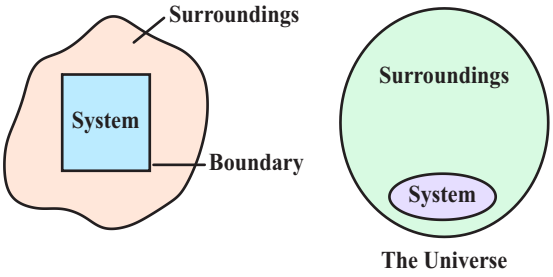


Figure 16.3 System, Boundary and Surrounding

Thermodynamic Systems:

1. **Open System:**
   In an open system, both *heat and matter can be exchanged* with the surroundings. For instance, steam (matter) can escape from the coffee, and heat can be lost to the surrounding air.
   Example: A cup of hot coffee.
2. **Closed System:**
   In a closed system, *only heat can be exchanged* with the surroundings, but matter cannot. In this example, heat can still transfer through the pot, but the water vapor (matter) is trapped inside the pot and cannot escape.
   Example: A pot with a lid.
3. **Isolated System:**
   In an isolated system, *neither heat nor matter can be exchanged with the surroundings*. A thermos flask is designed to prevent heat transfer, keeping the

contents inside hot or cold for a long period without any exchange with the environment.

Systems are as shown in figure 16.4 respectively

Example: A thermos flask.

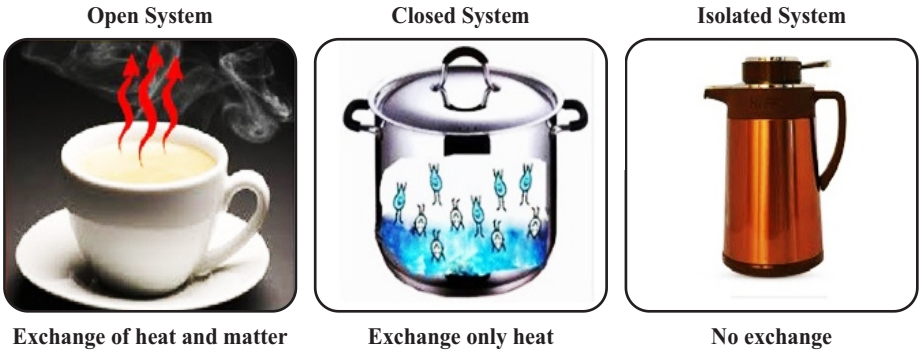


Figure 16.4 Open, Close and Isolated system

16.3.2 Work Done by a Gas:

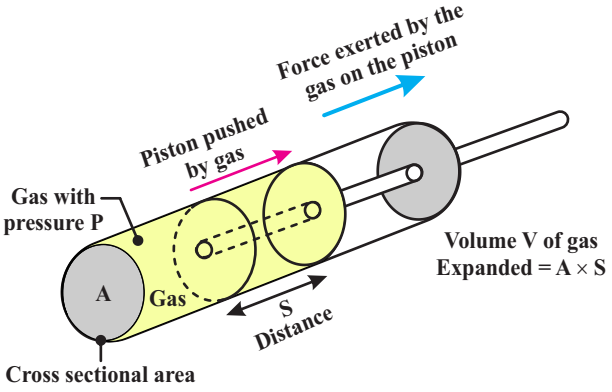
When a gas expands, it does work on its surroundings by exerting pressure on the walls of the container it's in. This is important, for example, in a steam engine where expanding steam pushes a piston to turn the engine. The work done when a volume of gas changes at constant pressure is defined as:

$$W = p\Delta V$$

Where:

- W = work done (J)
- p = external pressure (Pa)
- V = volume of gas (m^3)

For a gas inside a cylinder enclosed by a moveable piston, the force exerted by the gas pushes the piston outwards, Therefore, the gas does work on the piston



The gas expansion pushes the piston a distance S

Figure 16.5 Work Done by a Gas

Derivation:

The volume of gas is at constant pressure. This means the force F exerted by the gas on the piston is equal to:

$$F = p \times A$$

Where: p = pressure of the gas (Pa) and A = cross-sectional area of the cylinder (m^2)

The definition of work done is:

$$W = F \times S$$

Where: F = force (N) and S = displacement in the direction of force (m)

The displacement of the gas d multiplied by the cross-sectional area A is the increase in volume ΔV of the gas:

$$W = P \times A \times S \ldots\ldots 16.4$$

This gives the equation for the work done when the volume of a gas changes at constant pressure:

$$W = P\Delta V$$

Where: ΔV = increase in the volume of the gas in the piston when expanding (m^3)

This is assuming that the surrounding pressure p does not change as the gas expands, This will be true if the gas is expanding against the pressure of the atmosphere, which changes very slowly, When the gas expands (V increases), work is done by the gas. When the gas is compressed (V decreases), work is done on the gas.

Worked Example 16.1

The volume occupied by 1.00 mol of a liquid at 50 °C is 2.4×10^{-5} m^3. When the liquid is vaporized at an atmospheric pressure of 1.03×10^5Pa, the vapor has a volume of 5.9×10^{-2} m^3.The latent heat to vaporize 1.00 mol of this liquid at 50 °C at atmospheric pressure is 3.48×10^4 J. Determine change in internal energy ΔU of the system.

Solution:

Step 1: Write down the first law of thermodynamics

$\Delta U = Q - W$

Step 2: Write the value of heating q of the system

This is the latent heat, the heat required to vaporize the liquid = $\mathbf{3.48 \times 10^4\ J}$

Step 3: Calculate the work done W

$$W = P.\Delta V$$

$\Delta V = V_f - V_i = 5.9 \times 10^{-2} - 2.4 \times 10^{-5} = 0.058976\ m^3$

p = atmospheric pressure = 1.03×10^5 Pa

$W = (1.03 \times 10^5) \times 0.058976 = 6074.528 = 6.07 \times 10^3$ J

Since the gas is expanding, this work done is negative

$W = -6.07 \times 10^3$ J

Step 4: Substitute the values into first law of thermodynamics

$\Delta U = 3.48 \times 10^4 \quad (-6.07 \times 10^3) = 28\ 730 = 29\ 000$ J

$T_f = = \frac{9}{5} \times 36.88 + 32$

$T_f = 98.4°F$

Result: Result: $T_f = 98.4°$

Worked Example 16.2

When a balloon is inflated, its rubber walls push against the air around it. Calculate the work done when the balloon is blown up from 0.015 m^3 to 0.030 m^3. Atmospheric pressure = 1.0 × 10^5 Pa.

Solution:

Step 1: write the known quantities and point out quantities to be found

$V_1 = 0.015\ m^3$

$V_2 = 0.030\ m^3$

$P = 1.0 \times 10^5$ Pa.

W= ?

Step 2: Write down the equation for the work done by a gas

$$W = P\Delta V$$

Step 3: Now First, calculate ΔV:

$$\Delta V = V_f - V_i$$

$\Delta V = 0.030\ m^3 - 0.015\ m^3 = 0.015\ m^3$

Step 4: Now, substitute the values into the formula:

$$W = -P\Delta V$$

W = (1.0 × 105) × 0.015 = 1500 J

Result: The work done is **1500 J**

16.3.3: The first law of thermodynamics, internal energy and the heating of the system& conservation of energy:

First law of thermodynamics:

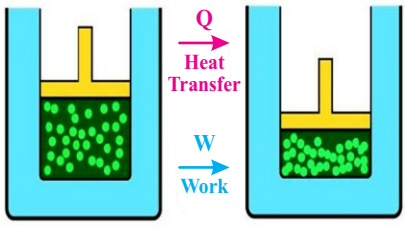
The first law of thermodynamics is based on the idea that energy can neither be created nor destroyed in any thermodynamic system. So, it is a particular form of the law of conservation of energy, which deals only with the heat energy.

"This law states that every thermodynamic system possesses a state variable (U) called the internal energy".

In any thermodynamic process, when heat energy (ΔQ) is added to a system, this energy appears as an increase in the internal energy (ΔU) stored in the system plus the work done (ΔW) by the system on its surroundings.

$$\Delta Q = \Delta U + \Delta W \ldots\ldots\ldots\ldots 16.5$$

ΔQ is taken positive when heat enters the system and negative when it leaves the system. ΔU is taken positive when the temperature of the system rises, while it is negative when the temperature of the system decreases.



ΔU = ΔQ – W

Figure 16.6 states of system

The first law of thermodynamics states that,

"The change in internal energy of a system equals the net heat transfer into the system minus the net work done by the system"

In equation form, the first law of thermodynamics is,

$$\Delta U = \Delta Q - W \ldots\ldots 16.6$$

ΔU →*change in internal energy* of the system

ΔQ → *net heat transferred into the system*

W →*net work done by the system*

The sum of all the kinetic (translational, rotational, and vibrational) and potential energies associated with the random motion of the atoms of a substance or system is the internal energy (U) of the substance.

According to the first law of thermodynamics Eq. No 16.6

$$\Delta U = \Delta Q - \Delta W$$

It means that the change in the internal energy of a system is equal to the energy flowing in as heat minus the energy flowing out as work. In other words, the change in the internal energy of the system is measured by the energy retained by the system. The energy absorbed by the system changes the translational, vibrational, and rotatio nal kinetic energy of the molecules. It also changes the potential energy of the molecules due to intermolecular forces.

upon the initial and final states of the system and not on the path taken between these two states. If the state of a thermodynamic system is changed from state (A) to state (B) then the change in internal energy of the system is

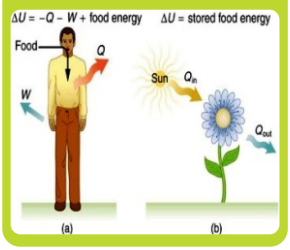
$$\Delta U = U_B - U_A \ldots\ldots\ldots 16.7$$

Putting this value in equation (16.6)
we have,

$$U_B - U_A = \Delta Q - \Delta W \ldots\ldots 16.8$$

DO YOU KNOW?

Examples of 1st law of thermodynamics Figure No.. (a) The first law of thermodynamics applied to metabolism. Heat transferred out of the body (Q) and work done by the body (W) remove internal energy, while food intake replaces it. (Food intake may be considered as work done on the body.) (b) Plants convert part of the radiant heat transfer in sunlight to stored chemical energy, a process called photosynthesis.



Applications of the first law thermodynamics:

Applications of first law of thermodynamics are all based on the principle that "the three forms of energy: internal energy; heat; and work can be inter converted."

The system is then often called the working substance. The mathematical formulation of the law $\Delta Q = \Delta U + \Delta W$ suggests the following processes that can be used for practical applications:

- The processes in which any one of the three terms of the equation is zero.
- The processes in which any one of the state variables P,V,T of the system are held constant.

(a) Isochoric Process:

"The thermodynamics process during which the volume of the system remains constant is called isochoric process".

We consider the gas contained in a cylinder having a conducting base and non-conducting walls and with a fixed piston at the end as shown in the Figure 16.7 Let heat ΔQ be imparted to the gas. The gas is then heated at constant volume. The pressure of the gas increases from P_1 to P_2 while its temperature increases from T_1 to T_2. Since the system neither expands nor contracts, work is neither done by the system nor on the system i.e ΔW=0.

Using the first law of thermodynamics equation (Eq. No 16.5)
we have:

$$\Delta Q = \Delta U + \Delta W$$
$$\Delta Q = \Delta U$$

This means that in an isochoric process the entire amount of heat supplied to the gas is converted to the internal energy of the gas. The pressure and temperature of the gas will increase. On the contrary, removal of heat from a system under isochoric condition will cause an equivalent decrease in the internal energy. The system will cool down and the pressure will fall.The graph of isochoric process is called an "isochor', which is a straight line, parallel to the pressure axis.

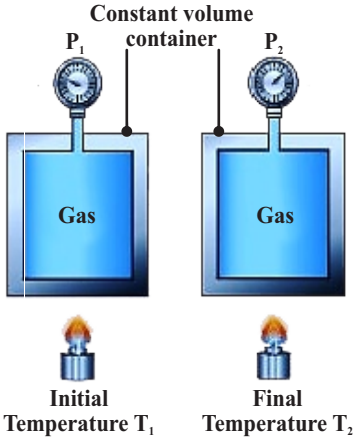


Figure 16.7 (a) Isochoric Process

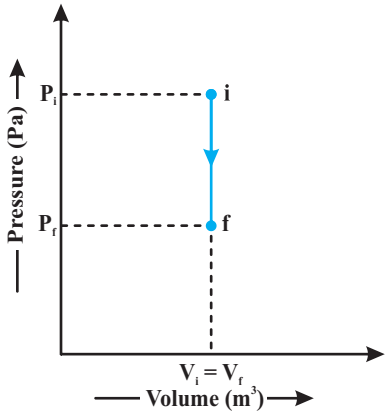


Figure 16.7 (b)

(b) Isobaric Process:

"The thermodynamics process during which the pressure is kept constant is called an isobaric process".

Isobaric expansion of a system is often used to convert heat into work. Practically all heat engines depend on the transformation of heat into work. Consider a gas contained in a cylinder having a conducting base and non-conducting walls and frictionless piston of cross-sectional area (A) as shown in figure 16.8(a).

When the gas is heated, a certain amount of heat energy is transferred into the system.

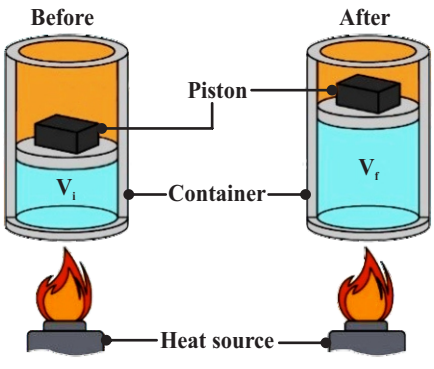


Figure 16.8(a) Isobaric Process

- The gas expands and moves the piston outward.
- Temperature changes from T_1 to T_2 and Volume changes from V_1 to V_2.
- If the displacement of the piston is kept very small, the pressure of the gas will not change much and can be considered constant.
- The work done by the gas which expands at constant pressure is:

$$\Delta W = P(V_2 - V_1)$$

$$\Delta W = P\Delta V$$

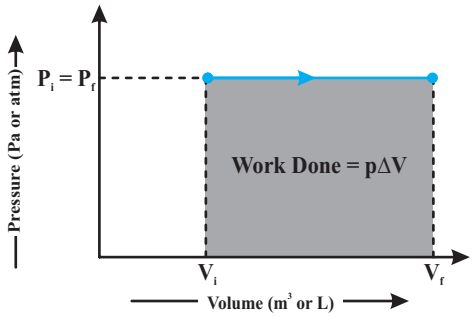


Figure 16.8 (b)
work done by change in volume of gas

Using the first law of thermodynamics equation

$$\Delta Q = \Delta U + \Delta W$$

$$\Delta Q = \Delta U + P\Delta V \ldots..16.9$$

The work performed by the expanding or contracting gas comes from one or both sources:

- heat supplied to the gas and
- the internal energy of the gas

The graph of isobaric process is called an "isobar", which is a straight line, parallel to the volume axis.

(c) Isothermal Process:

"The thermodynamics process which is carried out in such a way that a system undergoes changes but its temperature remains constant is called an isothermal process".

The internal energy of an ideal gas does not depend on the volume but depends only on the temperature of the gas. Consider a gas contained in a cylinder having a conducting base and non-conducting walls and with a movable piston at the end.

The base of the cylinder is placed on a heat reservoir at temperature T_1.Let the gas be then allowed to expand slowly by decreasing the pressure on the piston. Due to this expansion, the gas tends to cool down. But heat is conducted from the heat reservoir to the gas so that the temperature of the system remains constant and is equal to the temperature of the reservoir. Pressure changes from P_1 to P_2, volume changes from V_1 to V_2 but temperature remains constant (T).

During isothermal expansion some work is done by the gas in pushing up the piston in the cylinder. Since the temperature remains constant, there is no change in the internal energy of the gas, that is,$\Delta U = 0$

$$\Delta Q = \Delta U + \Delta W$$

$$\Delta Q = 0 + \Delta W$$

$$\Delta Q = \Delta W \ldots\ldots\ldots 16.10$$

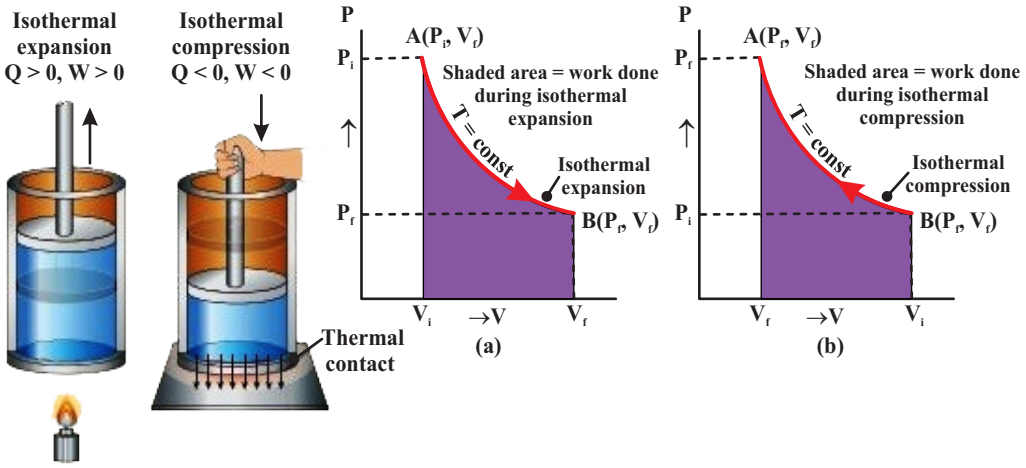


Figure 16.9 Isothermal Expansion, Compression and their graph (a, b)

This shows that:

If the gas expands and does external work, an equal amount of heat has to be supplied in order to maintain its temperature constant.

Conversely, if the gas contracts, work is being done on it and equal amount of heat has to be allowed to leave the gas. In the PV diagram the work done during the isothermal expansion is equal to the area under the graph as shown in Figure 16.9

Similarly, for an isothermal compression, the area under the PV graph is equal to the work done on the gas which turns out to be the area with a negative sign.

The whole process from an initial state to a final state is represented by a continuous curve which is called an "isotherm".

(d) Adiabatic Process:

"The thermodynamics process during which no heat enters or leaves a system is called an adiabatic process".

For every adiabatic process $\Delta Q = 0$, A truly adiabatic process is an ideal one which cannot be realized. However, the flow of heat may be prevented either by surrounding the system with a thick layer of heating insulating material such as cork, asbestos, or by performing the process very quickly. The flow of heat requires finite time, so any process performed quickly enough will be practically adiabatic. To perform an adiabatic process on a gas, we consider the gas to be contained in a completely insulated cylinder with a movable piston at the end.

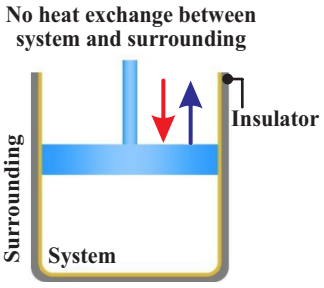


Figure 16.10 Adiabatic Process

Adiabatic Expansion:

If the gas is allowed to expand suddenly decreasing the pressure on the piston, the gas will be cooled, so that its temperature will decrease. Therefore, the internal energy of the gas will decrease.

Adiabatic Compression:

If the gas is compressed by suddenly increasing the pressure on the piston, the gas will be heated, so that temperature will increase. Therefore, the internal energy of the gas will increase.

According to the first law of thermodynamics;

$$\Delta Q = \Delta U + \Delta W$$
$$0 = \Delta U + \Delta W$$
$$\Delta U = -\Delta W$$

Thus, an increase in the internal energy of the system in an adiabatic process is equal to the work done on the system.

Above equation can be written as;

$$\Delta W = -\Delta U \ldots\ldots\ldots 16.11$$

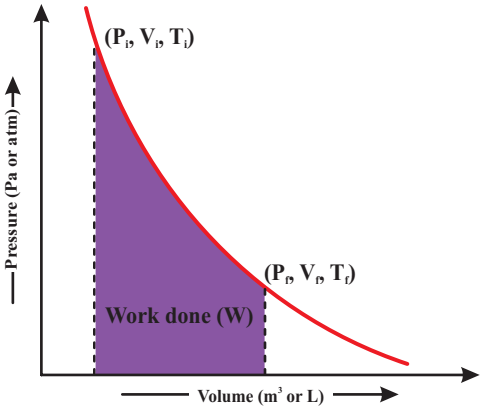


Figure 16.11 graph for adiabatic Compression

This means that if the system does the work, then in adiabatic process, the work is done at the cost of internal energy.

Self-Assessment Questions:

1. What is the first law of thermodynamics and how is it expressed mathematically?
2. How does the first law of thermodynamics relate to the conservation of energy?

Worked Example 16.3

The normal human body temperature is 36.88°C. What is this temperature on Fahrenheit scale? A gas confined in a cylinder undergoes an expansion from an initial volume of 2.0 L to a final volume of 4.0 L against a constant external pressure of 2.0 atm. Calculate the work done by the gas during this process.

Solution:

Step 1: Write the known quantities and point out quantities to be found.

Initial volume V_i= 2.0 L

Final volume V_f= 4.0 L

Constant external pressure (P) = 2.0 atm

Step 2: Write the formula and rearrange if necessary.

The work done by a gas against a constant external pressure can be calculated using the formula:

$$W = -P\Delta V$$

where: W is the work done,

P is the constant external pressure, and

ΔV is the change in volume.

Step 3: First, calculate ΔV:

$$\Delta V = V_f - V_i$$
$$\Delta V = 4.0L - 2.0L = 2.0L$$

Step 4: Now, substitute the values into the formula:

$$W = -P\Delta V$$
$$W = -(2.0atm) \times (2.0L)$$
$$W = -4.0atm \cdot L$$

Result: The negative sign indicates that work is done on the gas (compression) rather than by the gas (expansion). The final result is −4.0atm·L of work done on the gas during this expansion process.

Remember that the unit of work in the International System of Units (SI) is the joule (J), so you may want to convert the result to joules if needed (1 atm·L = 101.325 J).

Self-Assessment Questions:

1. How does an adiabatic process differ from an isothermal process in thermodynamics?
2. Why is an isothermal process important in thermodynamics, and how does it affect the internal energy of a system?
3. Why is an isobaric process significant in thermodynamics, and how does it affect the volume of a gas?

16.4.1 Heat capacity:

When heat is absorbed by a body, its temperature increases, and when heat is lost, its temperature decreases. Temperature represents the total kinetic energy of the particles comprising the body. Therefore, when heat is absorbed, it increases the kinetic energy of these particles, thereby raising the temperature. Consequently, the change in temperature ΔT is directly proportional to the amount of heat transferred q.

The formula $Q = nC\Delta T$ expresses the relationship where q is the heat transferred, n is the amount of substance (often expressed in moles), C is the molar heat capacity, and ΔT is the change in temperature. Molar heat capacity C refers to the amount of heat energy required to change the temperature of 1 mole of a substance by 1 unit. It is influenced by the substance's properties such as its nature, size, and composition.

Specific Heat:

Specific heat, often denoted by C orc, is a measure of the heat energy required to raise the temperature of a unit mass (usually one gram or one kilogram) of a substance by one degree Celsius (or one Kelvin). The formula for specific heat (c) is given by:

$$\boldsymbol{Q = mc\Delta T} \ldots\ldots\ldots \mathbf{16.12}$$

where:

Q is the heat energy absorbed or released,
m is the mass of the substance,
c is the specific heat, and
ΔT is the change in temperature.

Specific heat is an intensive property, meaning it does not depend on the quantity of the substance.

Molar Specific Heat:

Molar specific heat, denoted as C_m, measures how much heat energy is needed to raise the temperature of one mole of a substance by one degree Celsius or Kelvin. It's calculated using the formula:

$$Q = nC_m\Delta T \ldots\ldots..16.13$$

where:

Q is the heat transferred,
n is the number of moles,
C_m is the molar specific heat,
ΔT is the temperature change.

Molar specific heat is an intensive property and relates to specific heat C through the equation C_m = nC

In thermodynamics, both specific heat and molar specific heat are vital for quantifying how heat energy changes with temperature in substances. These values vary depending on the substance, reflecting each substance's unique ability to absorb or release heat.

There are two types of molar heat capacity:

C_P (molar heat capacity at constant pressure)
C_V (molar heat capacity at constant volume)

The relationship between CP and CV for an Ideal Gas:

From the equation $q = n\,C\,\Delta T$, we can derive:

At constant pressure $q_p = n\,C_p\Delta T$ This value corresponds to the change in enthalpy, expressed as: $q_P = n\,C_P\Delta T = \Delta H$

Similarly, at constant volume $Q_v = n\,C_v\Delta T$

This value corresponds to the change in internal energy, expressed as:

$$Q_v = n\,C_v\,\Delta T = \Delta U$$

For one mole of an ideal gas (n=1)

$$\Delta H = \Delta U + \Delta(PV) = \Delta U + \Delta(RT) = \Delta U + R\,\Delta T \ldots..16.14$$

Therefore, we have: $\Delta H = \Delta U + R\,\Delta T$

Substituting the expressions for

$$C_p\Delta T = C_v\,\Delta T + R\,\Delta T$$
$$C_p = C_p - C_v = R \ldots..16.15$$

This equation shows the relationship between the molar heat capacities at constant pressure C_P and constant volume C_V for an ideal gas, where R is the universal gas constant.

Worked Example 16.4

Calculate the molar specific heat capacity of a gas when 2 moles of the gas absorb 1500 J of heat energy, and its temperature increases by 25 degrees Celsius.

Solution:

Step1: write the known quantities and point out quantities to be found

Q = 1500J, n = 2mol, and ΔT = 25°C.

Step 2: Write the formula and rearrange if necessary

The heat transfer (Q) is given by:

$Q = n \cdot Cm \cdot \Delta T$

Let's rearrange the formula to solve for Cm:

$Cm = n \cdot \Delta TQ$

Step 3: Now, substitute the values into the formula:

$Cm = 1500J/2mol \cdot 25°C$

$Cm = 30J/mol \cdot °C$

Therefore, the molar specific heat capacity of the gas is 30J/mol·°C.

Result: 30J/mol·°C.

Worked Example 16.5

A sample of water with a mass of 200 grams is heated, and its temperature rises from 25°C to 45°C. Calculate the heat transferred to the water. Given the specific heat capacity of water (c) is 4.18 J/g°C.

Solution: Write the known quantities and point out quantities to be found

$m=200g$

$c=4.18J/g°C$

Initial temperature (T1) = 25°C

Final temperature (T2) = 45°C

Calculate the temperature change (ΔT): $\Delta T=T2-T1=45°C-25°C=20°C$

Step 2: Write the formula and rearrange if necessary

heat transfer equation $Q=mc\Delta T$ to calculate the heat transferred,

Step 3: Now, substitute the values into the formula:

Substitute the values: Q=(200g)×(4.18J/g°C)×(20°C)

Calculate the result: Q=8360J

Result: The heat transferred to the water is 8360 joules.

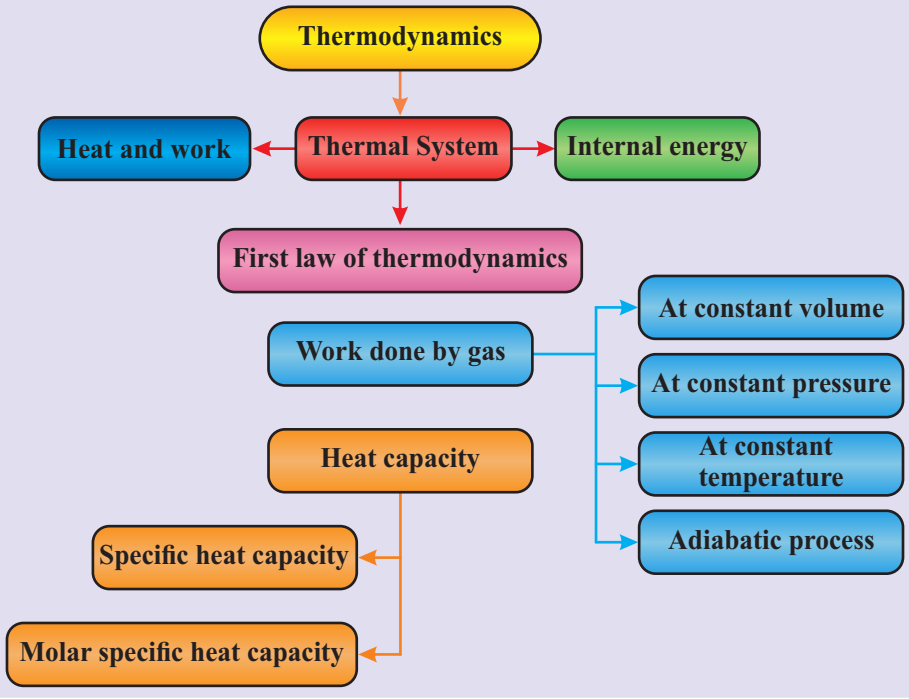
Self-Assessment Questions:

1. How does heat capacity relate to the amount of energy required to change the temperature of a substance?
2. Explain why different materials have different heat capacities.
3. How does the heat capacity of a substance change with temperature?

SUMMARY

- ✓ Heat (Q): Transfer of thermal energy between a system and its surroundings.
- ✓ Work (W): Mechanical energy transferred to or from a system through forces acting on it.
- ✓ Sign conventions: Heat added to the system and work done by the system are positive, while heat removed from the system and work done on the system are negative.
- ✓ Equation: $\Delta U = Q - W$, where ΔU is change in internal energy, Q is heat added to the system, and W is work done by the system.
- ✓ Internal energy (U): Total energy of a system, including kinetic and potential energies.
- ✓ First Law of Thermodynamics: Conservation of energy principle in thermodynamic systems.
- ✓ Energy cannot be created or destroyed, only transferred or converted from one form to another.
- ✓ Applications: Understanding energy transformations in various processes such as heating, cooling, phase changes, and chemical reactions.

CONCEPT MAP





Section (A): Multiple Choice Questions (MCQs)

Choose the correct answer:

1. What type of process occurs when a system exchanges both heat and work with its surroundings, and there is no change in internal energy?
   (a) Isothermal process (b) Adiabatic process
   (c) Isobaric process (d) Isochoric process
2. During an isobaric process, what remains constant?
   (a) Temperature (b) Pressure (c) Volume (d) Internal energy
3. In which thermodynamic process does a system exchange heat with its surroundings but undergoes no change in temperature?
   (a) Isothermal process (b) Adiabatic process
   (c) Isobaric process (d) Isochoric process
4. What is the characteristic of an adiabatic process?
   (a) Constant pressure (b) No heat exchange with the surroundings
   (c) Constant temperature (d) Constant volume
5. In an isochoric process, what is the primary feature?
   (a) Constant temperature (b) No work done
   (c) Constant volume (d) No heat exchange
6. What is internal energy in a thermodynamic system?
   (a) The energy associated with motion
   (b) The energy associated with the system's position
   (c) The sum of kinetic and potential energy
   (d) The total energy contained within the system
7. How is the change in internal energy (ΔU) defined in terms of heat (ΔQ and work (ΔW)
   (a) $\Delta U = \Delta Q + \Delta W$ (b) $\Delta U = \Delta Q - \Delta W$
   (c) $\Delta U = \Delta Q$ times W (d) $\Delta U = \Delta Q / \Delta W$
8. What is the internal energy of an ideal gas related to?
   (a) Temperature only (b) Pressure only
   (c) Volume only (d) Temperature, pressure, and volume
9. During an adiabatic process, what happens to the internal energy of the system?
   (a) Increases (b) Decreases
   (c) Remains constant (d) Depends on the specific heat
10. What is the equation for the internal energy change of a system in an isochoric process?
    (a) $\Delta U = \Delta Q + \Delta W$ (b) $\Delta U = \Delta Q - \Delta W$
    (c) $\Delta U = \Delta Q$ times ΔW (d) $\Delta U = \Delta Q / \Delta W$

Section (B): CRQs (Short Answered Questions):

1. Explain the concept of the first law of thermodynamics in your own words.
2. How does the first law of thermodynamics relate to the conservation of energy?

3. Distinguish between the work done by a system and the heat exchanged with the surroundings in the context of the first law.
4. Give an example daily life that illustrates the principles of the first law of thermodynamics.
5. Explain the role of the system and its surroundings in the context of the first law of thermodynamics.
6. How does heat capacity relate to the amount of energy required to change the temperature of a substance?

Section (C): ERQs (Long Answered Questions):

1. Provide the mathematical expression of the first law of thermodynamics and explain each term.
2. Describe what happens to the internal energy of a system in an adiabatic process, and why.
3. What is meant by the internal energy of a system, and how does it change during various thermodynamic processes?
4. Explore the concept of positive and negative work done by a gas, depending on whether the gas is expanding or compressing.
5. Explain the concept of heat capacity and its significance in thermodynamics. Discuss how heat capacity relates to the ability of a substance to store thermal energy and its implications for temperature changes.

Section (D): Numerical:

1. A gas undergoes isothermal expansion at a constant temperature of 300 K. If the gas absorbs 500 J of heat during the process, calculate the work done by the gas. **(500 J)**
2. A piston compresses a gas adiabatically. If the initial volume is 0.02 m^3 and the final volume is 0.01 m^3, and the initial pressure is 200 kPa, determine the final pressure. Assume the gas behaves ideally. **(400 kPa)**
3. A system undergoes an isobaric process where the pressure is kept constant at 150 kPa. If the volume increases from 0.05 m^3 to 0.08 m^3,calculate the heat added to the system. **(600 J)**
4. During an isochoric process, the internal energy of a gas increases by 300 J. If no work is done, determine the heat added to the system. **(300 J)**
5. A gas undergoes a cyclic process, starting at point A with a volume of 0.02 m^3, going to B (isochoric heating), then to C (isothermal expansion), and finally back to A. If the heat added during isothermal expansion is 1000 J and the heat rejected during isochoric heating is 500 J, calculate the net work done by the system. **(500 J)**
6. A gas expands from 0.03 m^3 to 0.06 m^3 against a constant pressure of 100 kPa. Calculate the work done in both a reversible and an irreversible process, and compare the results. **(300 J and 600 J)**
7. A 50 g piece of copper at 100°C is placed in 200 g of water at 20°C. If the final temperature of the system is 30°C, calculate the specific heat capacity of copper. (Specific heat capacity of water = 4.18 J/g°C) **(0.39 J/g°C)**
8. How much heat is required to raise the temperature of 1 kg of lead from 25°C to 100°C? (Specific heat capacity of lead = 0.128 J/g°C). **(0240J).**

Unit 17

Second Law of Thermodynamics

Teaching Periods 10 **Weightage %** 06



The neem tree, like all living systems, increases entropy in its surroundings through various processes, such as:

1. Transpiration: Releasing water vapors into the air, increasing humidity and entropy.

2. Photosynthesis: Absorbing sunlight and converting it into chemical energy, releasing excess energy as heat and increasing entropy.

These processes illustrate the second law of thermodynamics, where living systems increase entropy in their surroundings while maintaining their own organization and complexity.

In this unit student should be able to:

- State and explain second law of thermodynamics.
- State the working principle of heat engine.
- Describe the concept of reversible and irreversible processes.
- Describe the working of petrol engine and diesel engine.
- Explain the working principle of Carnot's engine
- Explain that the efficiency of a Carnot engine is independent of the nature of the working substance and depends on the temperatures of hot and cold reservoirs. Solve problems to find out the efficiency of heat engine
- Describe that refrigerator is a heat engine operating in reverse as that of an ideal heat engine and find its efficiency
- Solve problems to find out the efficiency of a refrigerator
- Describe that change in entropy is positive when heat is added and negative when heat is removed from the system.
- Explain that increase in entropy means degradation of energy.
- Show that energy is degraded during all natural processes.
- Identify that system tend to become less orderly over time
- Solve problems using the equation of entropy

Introduction:

The first law of thermodynamics is a particular form of the law of conservation of energy. However, there are some limitations regarding the transformation of heat energy into mechanical energy, which are covered by the 2nd law of thermodynamics.

17.1.1 Second law of thermodynamics:

Lord Kelvin and Rudolf Clausius formulated the second law of thermodynamics. It consists of two statements.

Kelvin Statement:

According to this statement ***"It is impossible to construct an engine, operating continuously in a cycle that can take heat from a source and coverts completely into work".***

According to Kelvin statement, an engine when converting heat into mechanical work cannot convert all of it into work. A part of heat must be rejected to a cooler reservoir, the exhaust as shown in figure 17.1(a).

In other words, it is impossible to device an engine that would have an efficiency of 100%, even though the first law of thermodynamics will be satisfied.

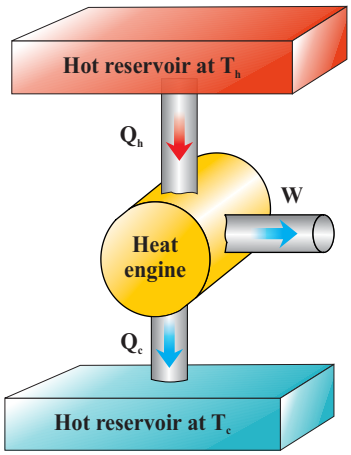


Figure 17.1 (a) the heat engine

Clausius Statement:

It is fact that no one has ever built a refrigerator which will work without a supply of energy. A refrigerator is essentially a machine for conveying heat from one body at a low temperature to another at a high temperature as shown in figure: 17.1(b). According to Clausius statement

"It is impossible to cause heat to flow from a cold body to a hot body without the expenditure of energy".

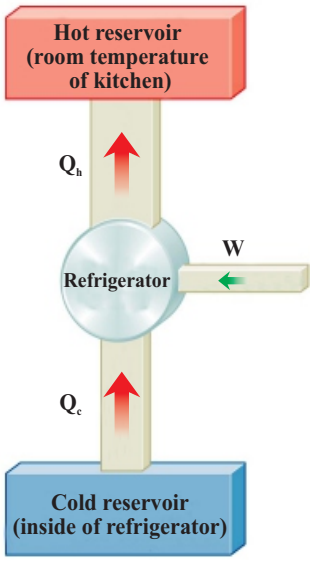


Figure 17.1 (b) the refrigerator

17.2.1 Working principle of heat engines:

Any device that transforms heat into mechanical energy (work) is called heat engine. The essentials of a heat engine are the furnace or hot body, the working substance and a condenser or cold body.

The simplest kind of engine to discuss is one in which the working substance undergoes a cyclic process, that is, a sequence of processes that eventually leaves the substance in the same state in which it started. In a steam turbine the water is recycled and used over and over. Internal combustion

engines do not use the same air again but we can still analyze them in terms of cyclic processes that approximate their actual operation.

All the heat engines absorbs heat from a source at a relatively low temperature performs some mechanical work and discards some heat at a low temperature the engine treats extra heat as waste. It gets rid of this waste heat through the exhaust pipes and the cooling system in internal combustion engines.

In a heat engine, Q_1 is positive but Q_2 is negative, representing heat leaving the working substance.

We can represent the energy transformation in a heat engine by the flow diagram fig: 17.2. The engine itself is represented by the circle. The amount of heat suppliedQ_1 to the engine by the hot reservoir is proportional to the cross section of the incoming pipeline at the top of the diagram. The cross section of the outgoing of the pipeline at the bottom is proportional to the magnitude $|Q_2|$of the heat discarded in the exhaust.

The branch line to the right represents the portion of the heat supplied that the engine converts to mechanical work W.

When an engine repeats the same cycle over and over, Q_1 and Q_2 represent the quantities of heat absorbed and rejected by the engine during one cycle. The net heat absorbed per cycle is

$$\Delta Q = Q_1 - Q_2$$

Let Q_1 be the heat absorbed by an engine from a high temperature reservoir and Q_2 be the heat rejected by the engine to low temperature reservoir or sink. The rest of heat energy is converted into useful work i.e.

$$W = Q_1 - Q_2 \ldots\ldots 17.1$$

DO YOU KNOW?

Each Q is positive when heat is transferred from a reservoir into the working substance and Q is negative when the reverse happens

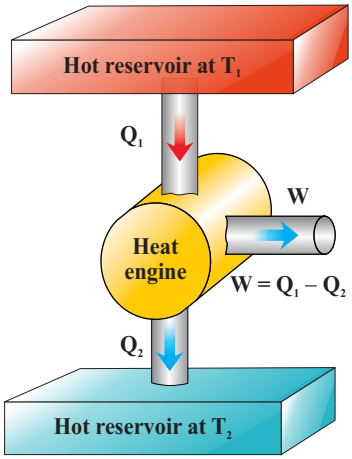


Fig: 17.2 Heat Engine

Efficiency of heat engine:

"***The thermal efficiency "η" of a cycle heat engine is defined to the ratio of the network 'W' done by the engine in each cycle to the heat absorbed Q_1 in each cycle".***

$$Efficiency\ (\eta) = \frac{output}{input} = \frac{W}{Q_1} = \frac{Q_1 - Q_2}{Q_1} = 1 - \frac{Q_2}{Q_1} \ldots\ldots 17.2$$

For 100% efficiency $Q_2 = 0$ i.e. no heat is rejected to low temperature reservoir (sink) and whole heat absorbed is converted into work. But experimental facts show that no cyclic engine can achieve an efficiency of 100%.

17.2.2 Reversible and Irreversible Processes:

To understand the nature of theoretical heat engine, we must first examine the concept of reversible and irreversible process. In a reversible process, the system undergoing the process

can be returned to its initial condition along the same path on a PV diagram and every point along this path is an equilibrium state. A process that does not satisfy these requirements is irreversible. All natural processes are known to be irreversible.

Consider an adiabatic process where a gas in a thermally insulated container as shown in fig: 17.3. A membrane separates the gas from a vacuum. When the membrane is burst, the gas expands freely into the vacuum. The system occupies a greater volume after the expansion. No heat energy is transferred to or from the container.

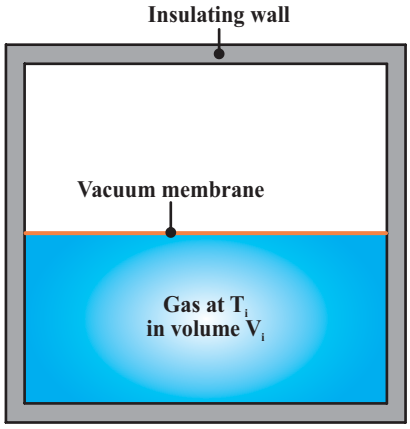


Figure 17.3 Thermally Insulated container

In this adiabatic process, the system undergoes changes without any heat exchange with its surroundings. Here is how the system changes. To make this process reversible, we must return the gas to its original volume and temperature without affecting the surroundings. If we try to reverse the process by compressing the gas back to its original volume using an engine, the surroundings change because external work is done on the system. The temperature increases during compression, which can be lowered by contact with an external energy reservoir. This returns the gas to its original state but affects the surroundings by adding energy. If this energy could drive the compression engine, the net energy transfer to the surroundings would be zero, making the process reversible. However, the Kelvin statement of the second law of thermodynamics specifies that the heat energy removed from the gas cannot be completely converted to mechanical work. Thus, the process is irreversible.

"A reversible process is one which can be retraced in exactly reverse order, without producing any change in the surroundings".

It is in equilibrium at all stages of operation. A succession of events which bring the system back to its initial condition is called a cycle. A reversible cycle is the one in which all the changes are reversible.

"The process which cannot be retraced in the backward direction by reversing the controlling factors is known as irreversible process"

It is in equilibrium only at initial and final stage. All changes which occur suddenly or which involved friction or dissipation of energy through conduction, convection or radiations are irreversible. An example of highly irreversible process is an explosion.

17.2.3 Petrol Engine:

Petrol engine is an internal combustion engine designed to run on volatile fuel such as petrol (gasoline), which has spark-ignition. In these engines air and fuel are generally mixed post-compression.

It works on the Otto cycle and the name comes from the German Engineer Nikolaus Otto (1876), who made the first working prototype. Although different engines may differ in

their construction technology but they are based on the principle of Carnot cycle. A typical four stroke petrol engine Fig. (17.4) also undergoes four successive processes in each cycle.

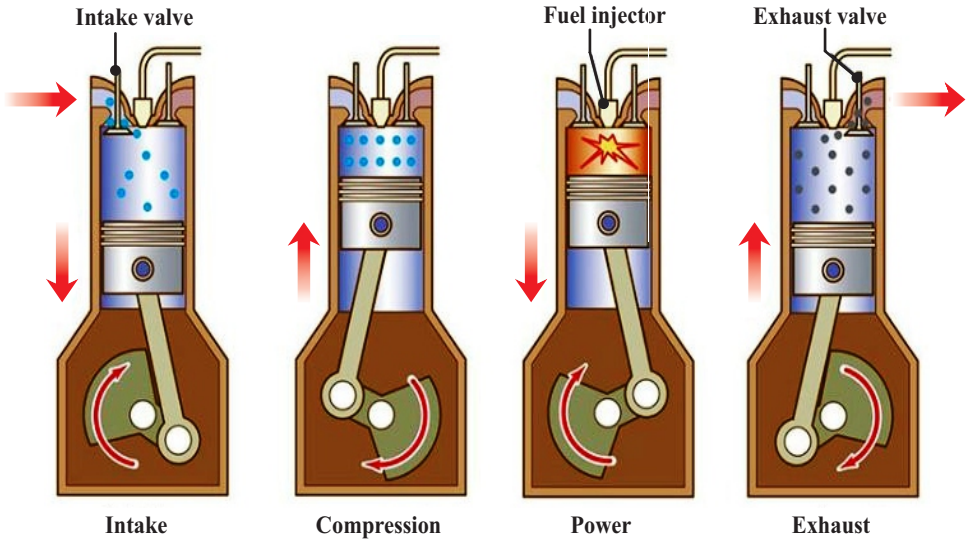


Figure17.4 Petrol Engine

1. The cycle starts on the intake stroke in which piston moves outward and petrol air mixture is drawn through an inlet valve into the cylinder from the carburetor at atmospheric pressure.
2. On the compression stroke, the inlet valve is closed and the mixture is compressed adiabatically by inward movement of the piston.
3. On the power stroke, a spark fires the mixture causing a rapid increase in pressure and temperature. The burning mixture expands adiabatically and forces the piston to move outward. This is the stroke which delivers power to crankshaft to drive the flywheels.
4. On the exhaust stroke, the outlet valves open. The residual gases are expelled and piston moves inward. The cycle then begins again. Most motorbikes have one cylinder engine but cars usually have four cylinders on the same crankshaft. The cylinders are timed to fire turn by turn in succession for a smooth running of the car. The actual efficiency of properly tuned engine is usually not more than 25% to 30% because of friction and other heat losses.

DO YOU KNOW?

No real engine operating between two energy reservoirs can be more efficient than a Carnot engine operating between the same two reservoirs in accordance with Carnot theorem.

17.2.4 Diesel Engine:

The diesel engine named after Rudolf Diesel (German inventor and Mechanical Engineer), is an internal combustion engine. He patented his original design in 1892. When the fuel comes into contact with high temperature, it ignites, creating energy that drives the piston

down transferring energy to the crankshaft. There are two classes of diesel engine: two strokes and four strokes. Most diesel engines generally use the four-stroke cycle.

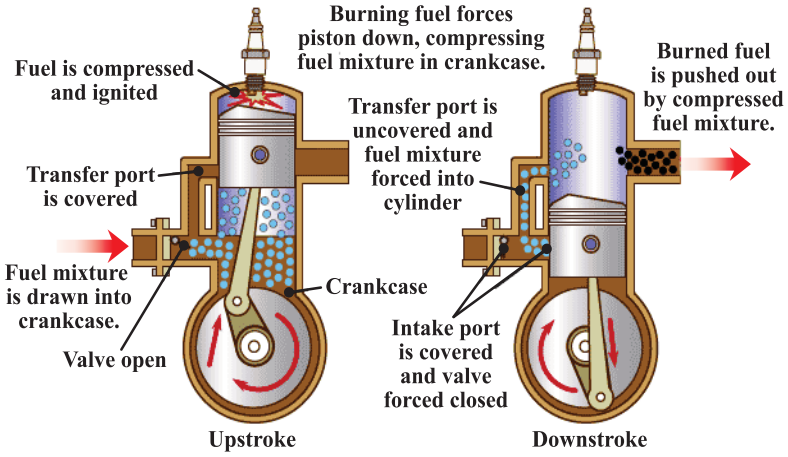


Fig: 17.5 Diesel Engine

No spark plug is needed in the diesel engine (fig: 17.5). Diesel is sprayed into the cylinder at maximum compression. Because air is at very high temperature immediately after compression, the fuel mixture ignites on contact with air in the cylinder and pushes the piston outward. The efficiency of diesel engine is about 35% to 40%.

17.3.1 The Carnot Engine:

A French Engineer named Sadi Carnot in 1824 described an ideal engine, now called a Carnot Engine, which is free from all sort of heat losses and friction. It consists of a hot body of infinite thermal capacity, a similar cold body, a perfect heat insulator and cylinder fitted with a piston enclosing any working substance. The cylinder has a heat conducting base and non-conducting walls and piston. The working substance is taken through the following cycle of operations known as the Carnot cycle. All real engines are less efficient than the Carnot engine because they do not operate through a reversible cycle. A cycle of a heat engine is completed when the properties of a system have returned to the original state. The operating cycle of the Carnot engine is called Carnot cycle. It consists of four processes, two isothermal and two adiabatic.

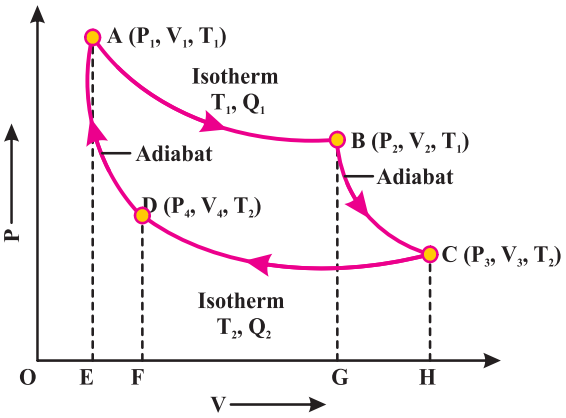


Figure 16.6 Carnot Cycle

1. In process A→B (fig.17.7 a) is isothermal expansion at temperature T_1. The cylinder is placed in thermal contact with an energy reservoir at temperature T_1.During the expansion, we decrease load on piston, the gas absorbs energy Q_1 from the reservoir through the base of the cylinder, volume of gas increases from V_1 to V_2 and does work W_{AB} in raising the piston.
2. In Process B→C (fig: 17.7 b), the base of the cylinder is replaced by a thermally non conducting wall and the gas expands adiabatically due to decrease in load on the piston; that is, no heat energy enters or leaves the system during the expansion, the temperature of the gas decreases from T_1 to T_2, volume increases from V_2 to V_3 and gas does work W_{BC} in raising the piston.
3. In process C→D (fig.17.7 c), the cylinder is placed in thermal contact with an energy reservoir at temperature T_2 is compressed isothermally at temperature T_2 by increasing the load on the piston. The volume decreases from V_3 to V_4. During this time, the gas expels energy Q_2 to the reservoir and the work done by the piston on the gas is W_{CD}.
4. In the final process D → A (fig. 17.7 d), the base of the cylinder is replaced by a non-conducting wall and the gas is compressed adiabatically by increasing the load on the piston. The temperature of the gas increases from T_2 to T_1 and volume decreases from V_4 to V_1 and work done by the piston on the gas is W_{DA}.

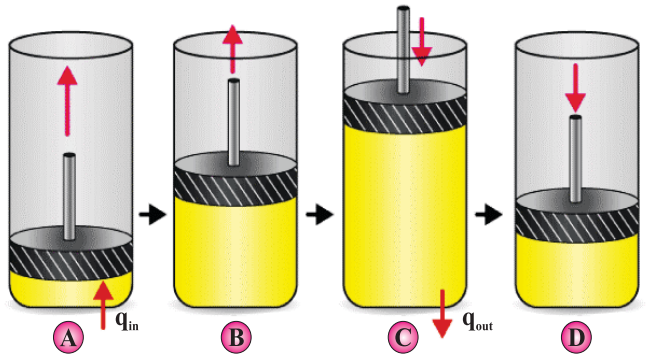


Fig: 17.7 (a,b,c,d) Carnot Engine cycle

If Q_1 is the heat absorbed by the working substance from the body in the isothermal expansion AB and Q_2 is the heat rejected to the cold body during the isothermal compression CD. In this cycle, the system comes back to its initial state and hence, there is no change in its internal energy. W is the external work done by the engine in one cycle; then from the 1st law of thermodynamics.

$$W = Q_1 - Q_2 \ldots\ldots\ldots\ldots\ldots 17.3$$

17.3.2 Efficiency of Carnot Engine:

Thermal efficiency (η) of a heat engine is the ratio of output to input.

$$Efficiency\ (\eta) = \frac{output}{input} = \frac{W}{Q_1} \ldots\ldots\ldots\ldots\ldots 17.4$$

From eq: 17.3, $W = Q_1 - Q_2$ then

$$\eta = \frac{Q_1 - Q_2}{Q_1} = 1 - \frac{Q_2}{Q_1} \ldots\ldots\ldots\ldots 17.5$$

Eq: 17.5 show that the efficiency of the engine increases as the ratio $\frac{Q2}{Q1}$ decreases. It can also be proved that the heat transferred to or from a Carnot engine is directly proportional to the temperature of the hot or cold body.

$$\frac{Q_2}{Q_1} = \frac{T_2}{T_1} \ldots\ldots\ldots\ldots 17.6$$

Thus, the efficiency of Carnot engine can be written as

$$\eta = 1 - \frac{T_2}{T_1} \ldots\ldots\ldots\ldots 17.7$$

The efficiency is usually taken in percentage, in that case.

$$\text{Percentage efficiency } (\eta) = \left(1 - \frac{T_2}{T_1}\right) \times 100\%$$

Thus, the efficiency of Carnot engine depends on the temperature of hot and cold reservoirs. The larger the temperature difference of two reservoirs the greater is the efficiency but it can never be one or 100% unless cold reservoir is at absolute zero temperature ($T_2 = 0^0K$)

Such reservoirs are not available and hence the maximum efficiency is always less than one. Nevertheless, the Carnot cycle establishes are upper limits on the efficiency of all heat engines.

No Practical heat engine can be perfectly reversible and also energy dissipation is inevitable. This fact is stated Carnot's theorem.

"No heat engine can be more efficient than a Carnot engine operating between the same two temperatures".

The Carnot's theorem can be extended to state that "all Carnot's engines operating between the same two temperatures have the same efficiency, irrespective of the nature of working substance". All real engines are less efficient than Carnot engine due to friction and other heat losses.

Worked Example 17.1

A Carnot's engine operates between two temperature reservoirs' of 227°C and 27°C, what is maximum efficiency of Carnot's engine?

Solution:

Step 1: Write down the known quantities and quantities to be found.

$T_1 = 227$ °C $\quad T_1 = 227 + 273 = 500K$ $\quad T_2 = 27$ °C $\quad T_2 = 27 + 273 = 300K$

Maximum efficiency = η =?

Step 2: Write down the formula and rearrange if necessary.

$\eta = 1 - \frac{T_2}{T_1}$

Step 3: Put the values and calculate

$$\eta = 1 - \frac{300}{500} = 1 - 0.6 = 0.4$$

$\eta = 0.4 \times 100\%$

$\eta = 40\ \%$

Result: 40 %

17.4.1 Refrigerator:

According to the 2nd law of the thermodynamics heat will always flow spontaneously from hot to cold body, and never the other way around.

A refrigerator causes heat to flow from cold to hot by doing work, which cools the space inside the refrigerator.

In a heat engine, the direction of energy transfer is from the hot reservoir to the cold reservoir, which is the natural direction. The role of heat engine is to process the energy from the hot reservoir so as to do useful work. What we wanted to transfer energy from the cold reservoir to the hot reservoir, because that is not the natural direction of energy transfer. We must put some energy into a device to be successful. Devices that perform this task are called refrigerator.

For example, houses in summer are cooled using refrigerators called air conditioners. The air conditioner transfers energy from the cool room in the home to the warm air outside.

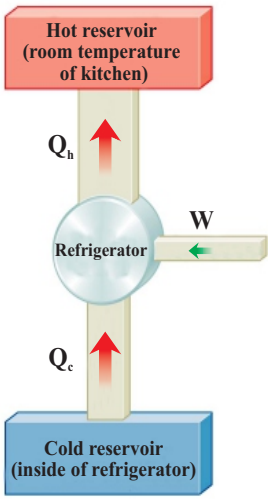


Fig: 17.8
Refrigerator or heat pump

In a refrigerator or heat pump, the engine takes in energy $|Q_C|$ from cold reservoir and expels energy $|Q_h|$ to a hot reservoir as shown in fig: 17.8, which can be accomplished only if work is done on the engine. From the first law of thermodynamics, we know that the energy given up to the hot reservoir must equal to sum of the work done and the energy taken in the form of the work done and the energy taken in the form of the cold reservoir.

$$Q_h = W + Q_c \ldots\ldots\ldots\ldots 17.8$$

Therefore, the refrigerator transfers energy from a colder body to a hotter body.

In Practice, a refrigerator includes a circulating fluid that gases through two sets of metal coils that can exchange energy with the surroundings. The fluid is cold and at low pressure when it is in the coils located in a cool environment, where it absorbs energy by heat. The resulting warm fluid is then compressed and enters the other coils as a hot, high-pressure fluid. There it releases its stored energy to the warm surroundings. In an air conditioner, energy is absorbed into the fluid in coils located in a building's interior; after the fluid is compressed, energy leaves the fluid through coil. In a refrigerator, the external coils are behind or underneath the unit Fig: 17.9 the internal coils are in the walls of the refrigerator and absorb energy from the food.

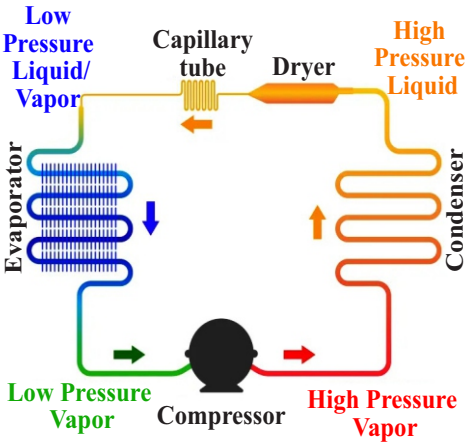


Figure 17.9 Refrigeration cycle

Efficiency of a Refrigerator:

The effectiveness of a refrigerator is described in terms of a number called the coefficient of performance (COP).

The COP is similar to the thermal efficiency for a heat engine in that it is a ratio of what you gain (energy transferred to or from a reservoir) to what you give (work input). For a refrigerator operating in the cooling mode, "what you gain" energy is removed from the cold reservoir the most effective refrigeration or air conditioner is one that removes the greatest amount of energy from the cooler reservoir in exchanges for the least amount of work. Therefore, for these devices operating in the cooling mode, we define the COP in terms of $|Q_C|$. The Coefficient of performance of a refrigerator is defined as

"The heat removed from the cold reservoir Q_C (i.e. inside refrigerator) divided by the work done to remove the heat (i.e. the work done by the compressor)".

$$\text{COP of refrigerator} = \frac{heat\ extracted}{work}$$

$$\text{COP} = \text{K} = \frac{Q_c}{w}$$

$$\therefore Q_h = Q_c + w$$

$$w = Q_h - Q_c$$

$$COP = K = \frac{Q_c}{Q_h - Q_c} \ldots\ldots\ldots\ldots 17.9$$

We know that for Carnot cycle $Q_1 \propto T_1$ and $Q_2 \propto T_2$

$$\therefore COP = K = \frac{T_c}{T_h - T_c} \ldots\ldots\ldots\ldots 17.10$$

T_c is the cryogenic temperature at which the heat is removed

T_h is the temperature at which the heat is rejected

It can also be seen from above relations that the coefficient of performance of refrigerator can be larger than 100% unlike the efficiency of heat engine which is always less than 100%. A good refrigerator should have a high COP, typically 5 or 6.

Worked Example 17.2

A certain refrigerator has a COP of 5.00. When the refrigerator is running, its power input is 500 watts. A sample of water of mass 500g and temperature 20.0°C is placed in the freezer component. How long does it take to freeze the water to ice at 0°C? Assume all other parts of refrigerator stay at the same temperature and there is no leakage of energy from the exterior, so the operation of the refrigerator result only is energy being extracted from the water.

Solution:

Step 1: Write down the known quantities and quantities to be found.

COP = K = 5.00,

P = 500 watt

Mass of water = m = 500g = 0.50 kg,

$T_1 = 20°C$

$T_2 = 0°C$,

$\Delta T = T_2 - T_1 = -20°C$

Specific heat of water = c = 4186 J/kg.°C

Latent heat of fusion $= L_f = 3.36 \times 10^5$ J/kg

Time = t =?

Step 2: Write down the formula and rearrange if necessary.

$$|Qc| = mc\Delta T - mL_f = m(c\Delta T - L_f)$$

$$= 0.50\,kg\,[(4186 J/kg.°C)(-20°C) - 3.36 \times 10^5\,J/kg]$$

$$= 2.08 \times 10^5 J$$

$$COP = \frac{|Q_c|}{w} \rightarrow w = \frac{|Q_c|}{COP} = \frac{2.08 \times 10^5}{5.00}$$

$$W = 4.17 \times 10^4 J$$

$$P = \frac{w}{t} \rightarrow t = \frac{w}{P} = \frac{4.17 \times 10^4\,J}{500w} = 83.4\,sec$$

Result: In reality the time interval for the water to freeze in a refrigerator is much longer than 83.4 s, which suggests that the assumptions of our model not valid. Only a small part of the energy extracted from the refrigerator interior in a given time interval comes from the water. Energy must also be extracted from the container in which the water is placed, and energy that continuously leak into the interior from the exterior must be extracted.

Worked Example 17.3

What is the coefficient of performance (COP) of a refrigerator that operates with Carnot efficiency between temperatures -3.00°C and +27.0°C?

Solution:

Step 1: Write down the known quantities and quantities to be found.

Tc = -3.00°C = 270K

Th = 27.0°C = 300K

COP = K =?

Step 2: Write down the formula and rearrange if necessary.

$$COP = K = \frac{Tc}{Th - Tc}$$

$$COP = K = \frac{270}{300-270} = \frac{270}{30} = 9$$

Result: The coefficient of performance of a refrigerator is 9.

Self-Assessment Questions:

1. An inventor claims that he has construct an engine which give efficiency of 56%. When working between -10 and 900, check whether his claim is true or false.
2. Why is the coefficient of performance (COP) used instead of efficiency in refrigerators?

17.5.1 Entropy:

Entropy was first expressed by German physicist and mathematician, Rudolf Clausius in 1865 into the study of thermodynamics to give a quantitative basis for the 2nd law of thermodynamics. The disorderness of the system is known as entropy. Entropy is also defined as

"The measure of a system's thermal energy per unit temperature that is unavailable for doing useful work".

We have discussed several processes that proceed naturally in the direction of increasing disorder. Irreversible heat flow increases disorder because the molecules are initially sorted into hotter and cooler regions; this sorting is lost when the system comes to thermal equilibrium. Adding heat to a body increases its disorder because it increases average molecular speed and therefore the randomness of molecular motion. Free expansion of a gas increases its disorder because the molecules have more randomness of position after the expansion than before.

Entropy provides a quantitative measure of disorder. T_0 introduce the concept, let's consider an isothermal expansion of an ideal gas; we add heat Q and let the gas expand just enough that the temperature remains constant. Because the internal energy of an ideal gas depends only on its temperature, the internal energy is also constant, so from the first law of thermodynamics the work w done by the gas during this expansion is equal to the heat Q added to it. That is

$$\Delta Q = \Delta W = P.\Delta V$$

The gas is in a more disordered state after the expansion than before, because the molecules are moving in a large volume and have more randomness of position. Thus ΔV is a measure of the increase in disorder of the molecules, and we see that it is proportional to the quantity $\frac{\Delta Q}{T}$. We introduce the symbol S, called the entropy of the system, and we define the entropychange ΔS during a reversible isothermal process as

$$\Delta S = \frac{\Delta Q}{T} \ldots\ldots\ldots\ldots 17.11$$

Change in entropy is positive when heat is added and negative when heat is removed from the system. Suppose an amount of heat Q flows from a reservoir at temperature T_1through a conducting rod to a reservoir at temperature when $T_1 > T_2$. The change in entropy of the reservoir at temperature T_1, which loses heat, decreases by $\frac{Q}{T_1}$, and of the reservoir at temperature T_2 which gains heat increases by $\frac{Q}{T_2}$. As $T_1 > T_2$ so $\frac{Q}{T_2}$ will be greater than $\frac{Q}{T_1}$. i.e. $\frac{Q}{T_2} > \frac{Q}{T_1}$. Hence, net change in entropy = $\Delta S = \frac{Q}{T_2} - \frac{Q}{T_1}$ is positive. It follows that in all natural processes where heat flows from one system to another; there is always a net increase in entropy. This is another statement of 2nd law of thermodynamics.

"The entropy of the universe during any process either remains constant or increases"

Similarly, we can prove that in every natural process the entropy always increases provided we considered the whole system. It seems to be a law of nature that a natural process always takes place in such a direction as to cause an increase in the entropy of the system and its surroundings. It is known as the law of increase of entropy.

The S.I unit of entropy is JK^{-1}. If heat is removed from a system, the change in entropy is written as negative.

17.5.2 Increase in Entropy Means Degradation of Energy:

Suppose a quantity of heat Q in a reservoir at temperature T_1. Let the temperature of the coldest available reservoir be T_o. A Carnot engine working between the temperatures T_1 and T_o can absorb heat Q at temperature T_1 and do useful work W_1 given by:

$$W_1 = Q - Q_o$$

The efficiency of Carnot engine is

$$\eta = \frac{W_1}{Q} = 1 - \frac{T_o}{T_1}$$

$$W_1 = Q\left(1 - \frac{T_o}{T_1}\right) \ldots\ldots\ldots\ldots 17.12$$

Equation (17.12) gives maximum available energy which can be converted into useful work, when heat Q is stored in the reservoir at temperature T_1.

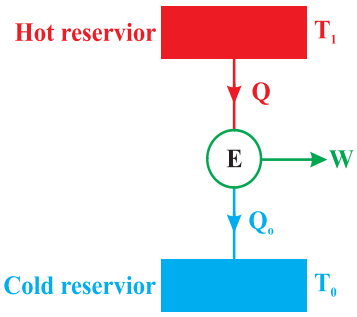


Figure:17.10 (a) Entropy

Consider an irreversible process, in which heat Q flows from the reservoir at temperature T_1. to another reservoir at a lower temperature T_2. A Carnot engine working between the temperatures T_2. and To can now take heat Q from the reservoir at temperature T_2. and do useful work W_2given by:

$$Efficiency = \eta = \frac{W_2}{Q} = 1 - \frac{T_o}{T_2}$$

$$W_2 = Q\left(1 - \frac{T_o}{T_2}\right) \ldots\ldots\ldots 17.13$$

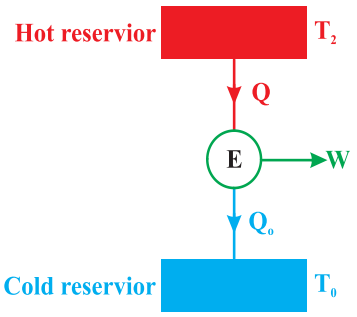


Figure:17.10 (b) Entropy

Equation (17.13) gives the maximum available energy which can be converted into useful work. When heat Q is stored in the reservoir at lower temperature T_2. As $T_2 < T_1$, we see from equation (17.12) and (17.13) that W_2 is less than W_1, i.e. available energy decreases with the increase of entropy during an irreversible process. Since all natural processes are irreversible, we conclude that the energy of the universe is continuously becoming unavailable for useful work. This is called the degradation of energy. From the equation (17.12) and (17.13), we get

$$W_1 - W_2 = Q\left(\frac{T_o}{T_2} - \frac{T_o}{T_1}\right)$$

$$W_1 - W_2 = \left(\frac{Q}{T_2} - \frac{Q}{T_1}\right) T_o \ldots\ldots\ldots 17.14$$

Now $\frac{Q}{T_1}$ = decrease in entropy of the reservoir at temperature T_1 and $\frac{Q}{T_2}$ = increase in entropy of the reservoir at temperature T_2.

Thus $\left(\frac{Q}{T_2} - \frac{Q}{T_1}\right)$ is increase in entropy of the universe. Also $(W_1 - W_2)$ is the amount of energy which has been degraded or made available for useful work. Hence equation (17.14)

shows that the increase of unavailable energy is equal to the increase in entropy of the universe multiplied by the temperature of the coldest available reservoir. Using eq: (17.14) an entropy increases, so unavailable energy also increases, so the useful energy decreases. It is called degradation of energy.

17.5.3 Energy is Degraded During All-Natural Processes:

It is believed that the temperature of the universe is increasing gradually and its entropy is also increasing. The mode of increase of entropy indicates that after a very long time, the entropy of the universe will be maximum and all the objects will be at the same temperature. At that time, no useful energy will be available and life will cease to exist, which is known as the "Heat Death" of the universe, which may not occur in near future.

It seems to be a law of nature that all natural processes always take place in such a direction so as to cause an increase in the entropy of a system and its surroundings.

Thus the 2nd law of thermodynamics is also expressed as the "Law of increase of entropy".

It states that, ***"when all the systems taking part in a process are included, the entropy either remains constant or increases".***

Symbolically, $\Delta S \geq 0$

The 2nd law of thermodynamics reveals that all the natural processes proceed towards a state of greater entropy, i.e. greater disorder or greater unavailability.

In all natural processes, energy always tends to pass from a more available (useful) form to a less available (degradation) form unavailability is a measure of entropy. A reversible process does not change total entropy of the universe. Thus $\Delta S = 0$; any irreversible process increases the total entropy of the universe $\Delta S > 0$, Total entropy change in one cycle of any Carnot engine is zero. All natural processes are irreversible and therefore, during all-natural process's energy is degraded.

17.5.3 Systems Tends to Become Less Orderly Over Time:

The disorder of the system increases during any natural process, a stage will reach when universe will approach a stage of maximum disorder. At this stage, matter will become a uniform mixture i.e. whole universe will be at uniform temperature. No work can then be done and heat flow will cease.

Time arrow concept of entropy tells us in which direction the time is going. It is the change in entropy that ultimately provides us with the answer to why systems will naturally evolve in one direction with time and not the other. Systems always evolve in time in such a way that the total entropy of system plus environment increases. If you observe a system in which the entropy appears to decrease, you can be sure that somewhere there is change in the entropy of the environment large enough to make the total entropy change positive. Entropy has been called a time arrow: i.e. events occur in the direction of increasing disorder with time.

In all the processes going on in practical life, the entropy always increases with time and disorder is produced more and more. Hence this increase of entropy is identification for the time passage.

Worked Example 17.4

What is the change in entropy of 30gm of water at 0°C as it is changed into ice at 0°C? Take latent heat of fusion of ice 336000 J kg^{-1}.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Change in entropy = ΔS =?

Mass of water = m =30gm = 0.03kg

T = 0°C = 0 + 273 = 273K

Specific latent heat of fusion of ice = H_f = 336000J kg^{-1}

Step 2: Write down the formula and rearrange if necessary.

$\Delta S = \frac{\Delta Q}{T}$

∴ Heat required freezing 30gm of water at 0°C to ice at 0°C.

$\Delta Q = mH_f = 0.03 \times 336000 = 10800J$

Put the values and calculate

$\Delta S = \frac{10080\ J}{273\ K} = 36.92\ Jk^{-1}$

Result: The change in entropy comes out 36.92 Jk^{-1}.

Worked Example 17.5

Calculate change in entropy when 10kg of water is heated from 90°C to 100°C. Specific heat of water = 4180J.$kg^{-1}K^{-1}$.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Change in entropy = ΔS =?

Mass of water = m =10kg

T_1 = 90°C

T_2 = 100°C

ΔT = 100 – 90 = 10°C

Average temperature = $T = \frac{T_1+T_2}{2} = \frac{90+100}{2} = 95°C$

T = 95°C + 273 = 368K

Step 2: Write down the formula and rearrange if necessary.

$\Delta S = \frac{\Delta Q}{T} = \frac{mc\Delta T}{T}$

Put the values and calculate

$\Delta S = \frac{10 \times 4180 \times 10}{368} = 1135.87\ Jk^{-1}$

Result: The change in entropy comes out 1135.87 Jk^{-1}

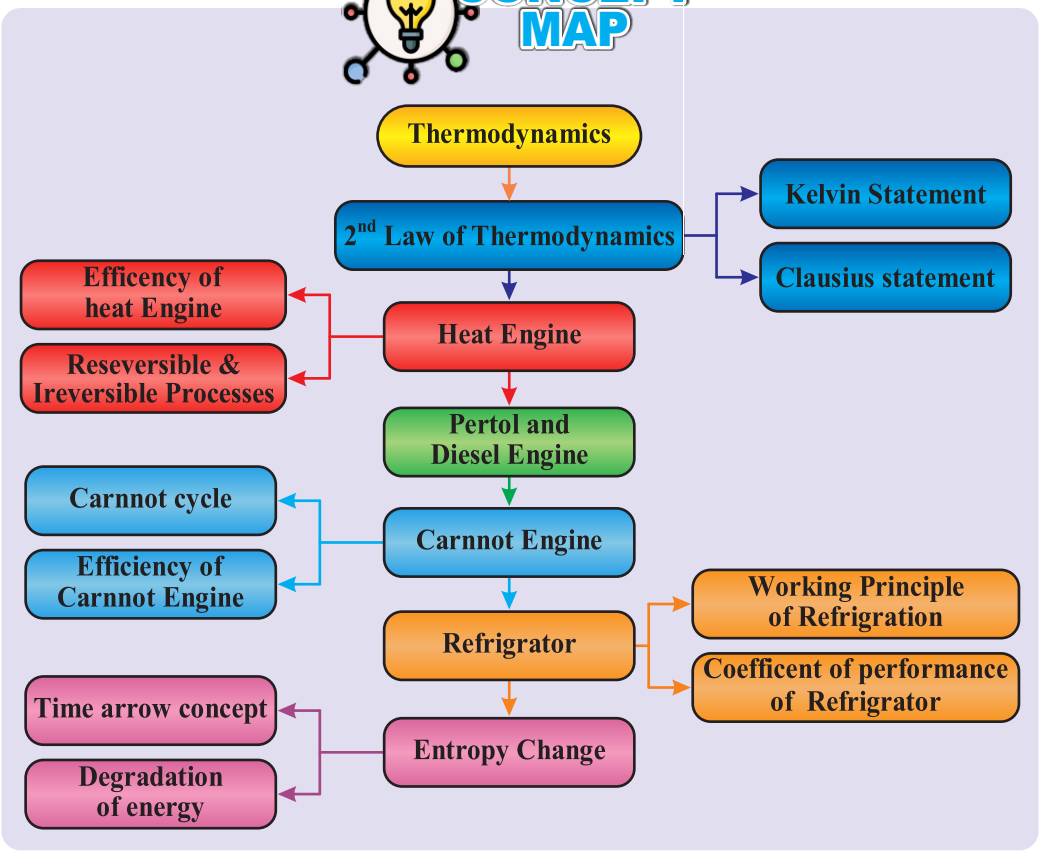


- ✓ Second law of thermodynamics can be states as:
  - ➢ There is no perpetual motion machine that can convert the given amount of heat completely into work.
  - ➢ The total entropy of any system plus that of its environment increases as a result of any natural process.
- ✓ A heat engine is a device which converts a part of thermal energy into useful work.
- ✓ Reversible Process: A process in which system and its surroundings can be returned to their initial conditions.
- ✓ Irreversible Process: A process which cannot be retraced in the backward direction by reversing the controlling factors.
- ✓ Efficiency of Carnot Engine is the ratio of output to input temperature and is $1 - \frac{T_2}{T_1}$
- ✓ Refrigerator is a heat engine operating in reverse as that of an ideal engine.
- ✓ Coefficient of performance of refrigerator (COP) is equal to the ratio of the heat taken from the cold reservoir to the work done by the refrigerator.

$$\text{COP} = \frac{T_c}{T_h - T_c}$$

- ✓ T_c is the cryogenic temperature at which the heat is removed and T_h is the temperature at which the heat is rejected.
- ✓ Entropy:
  - ➢ It is a measure of disorder (microscopic randomness) of a system.
  - ➢ It is a measure of unavailable energy of a system.
  - ➢ In terms of 2nd law of thermodynamics, "Every process in nature takes place such that the entropy of an isolated system either remains unchanged or it increases". $\Delta S \geq 0$.
- ✓ Entropy change ΔS due to heat transfer at absolute temperature T is given by $\Delta S = \pm \frac{\Delta Q}{T}$
- ✓ The S.I units of entropy are $J.K^{-1}$

CONCEPT MAP



EXERCISE

Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. A frictionless heat engine can be 100% efficient only if its exhaust temperature is:
   (a) 0°C (b) Equal to its input temperature
   (c) 0 K (d) half of its input temperature
2. A refrigerator, with its door open. The temperature of the room will
   (a) Rise (b) Fall
   (c) Remains the same (d) Rise or fall depending on the area of the room
3. Which device is not used in a diesel engine
   (a) Outlet Valve (b) Piston
   (c) Sparking Plug (d) Injector

4. Which one of the following processes is irreversible?
   (a) Slow compression of an elastic spring
   (b) Slow evaporation of a substance in an isolated vessel
   (c) Slow compression of a gas
   (d) A chemical explosion
5. According to 2nd law of thermodynamics, 100% conversion of heat into mechanical work is:
   (a) Possible (b) Possible, if the conditions are ideal
   (c) Not possible (d) Possible, if the process is adiabatic
6. The change in entropy is given by
   (a) $\Delta S = \frac{\Delta Q}{T}$ (b) $\Delta S = \Delta Q . T$ (c) $\Delta S = \frac{\Delta U}{T}$ (d) $\Delta S = \frac{\Delta W}{T}$
7. The net change in entropy as a system in a natural process is
   (a) Positive (b) Negative (c) Zero (d) Infinite
8. The efficiency of diesel engine is
   (a) Greater than petrol engine (b) Less than petrol engine
   (c) Equal to petrol engine (d) both have efficiency 1
9. Second law of thermodynamics states that:
   (a) Energy can't be converted. (b) Entropy decreases over time.
   (c) Heat flows from cold to hot. (d) Entropy increases over time
10. The process violates the 2nd law of thermodynamics is:
    (a) Refrigerator cooling. (b) Heat engine working.
    (c) Gases mixing. (d) Heat flowing from cold to hot.

Section (B): CRQs (Short Answered Questions):

1. What are some factors that affect the efficiency of automobile engines?
2. What happens to the temperature of a room in which an air conditioner is left running on table in the middle of the room?
3. Under what conditions can heat be added to a system without changing its temperature?
4. Is it possible to cool a room by keeping the refrigerator door open?
5. When does the entropy of a system decrease?
6. Is it possible, according to the 2nd law of thermodynamics to construct an engine that is free from thermal pollution?
7. Can heat be completely converted to work?
8. Define that why entropy has often been called as “time arrow”.

Section (C): ERQs (Long Answered Questions):

1. Give the two statements of the second law of the thermodynamics. Elaborate the concept of entropy and state the second law of thermodynamics in terms of this concept.
2. Explain the working principle of heat engine and also derive the formula for its efficiency.
3. Describe the concept of reversible and irreversible process.
4. What is Carnot engine? Give the operation of Carnot cycle and show that the efficiency of even this engine is less than 100%.

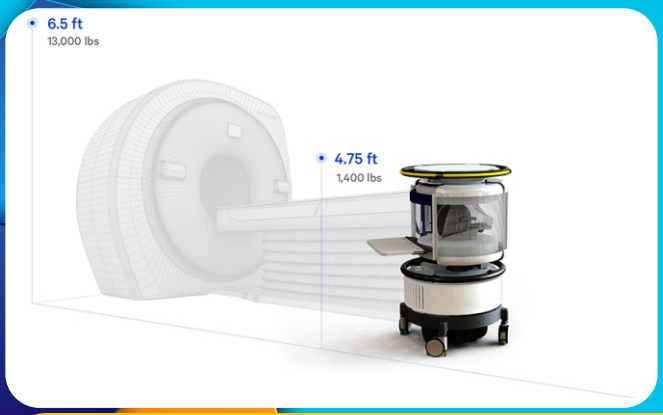
5. Describe that refrigerator is a heat engine operating in reverse as that of an ideal heat engine and find its efficiency.
6. Explain that change in entropy is positive when heat is added and negative when heat is removed from the system.
7. Explain that increase in entropy means degradation of energy.

Section (D): Numerical:

1. A Carnot engine takes 2000J of heat from a reservoir at 500K does some work, and discards some heat to a reservoir at 350K. How much heat is discarded, how much work does the engine do, and what is the efficiency? **(Ans: 1400J, 600J, 30%)**
2. One kilogram of ice at 0°C is melted and converted to water at 0°C. Compute its change in entropy. **(Ans: 1220 $J.K^{-1}$)**
3. In a high-pressure steam turbine engine, the steam is heated to 600°C and exhausted at about 90°C. What is the highest possible efficiency of any engine that operates between these two temperatures? **(Ans: 58.4 %)**
4. Temperature difference between the surface water and bottom water in Manchester Lake might be 5°C.Assuming the surface water to be at 20°C. What highest efficiency a steam engine could have if it operates between these two temperatures? **(Ans: 1.71%)**
5. A heat engine works at the rate of 500kW. The efficiency of the engine is 30%. Calculate the loss of heat per hour. **(Ans: 4.2 x 10^9J)**
6. A heat engine performs work of 0.4166 watts in one hour and rejects 4500J of heat to the sink. What is the efficiency of engine? **(Ans: 24.9%)**
7. A Carnot engine operates between the temperatures 850K and 300K. the engine performs 1200J of work in each cycle, which takes 0.25 sec
   (a) What is the efficiency of this engine?
   (b) What is the average power of this engine?
   (c) How much energy is extracted as heat from the high temperature reservoir?
   (d) How much energy is delivered as heat to the low temperature reservoir?
   [Ans (a) 65% (b) 4.8KW (c) 1855J (d) 655J]
8. A Carnot engine absorbs 52kJ as heat and exhaust 36kJ as heat in each cycle. Calculate:
   (a) The engine efficiency
   (b) The work done per cycle in kilojoules. **[Ans (a) 30.76% (b) 16KJ]**

Unit 18 Magnetic Fields

Teaching Periods 13 **Weightage %** 09



In 2021, Pakistan has become the first country in Asia and only the third outside the USA to acquire a portable MRI (magnetic resonance imaging) system in Karachi Sindh that can be wheeled directly to a patient's bedside and make MR imaging. MRI Machines uses strong magnetic fields to create detailed images of the inside of the human body.

In this unit student should be able to:

- Magnetic field is an example of a field of force produced either by current-carrying conductors or by permanent magnets.
- Describe and sketch field lines pattern due to a long current carrying straight wire.
- Describe the factors affecting the force on a current carrying conductor in a magnetic field.
- Solve problems using the equation F = BILsinθ, with directions as interpreted by Fleming's left-hand rule.
- Define magnetic flux density and the Tesla.
- Understand how the force on a current-carrying conductor can be used to measure the flux density of a magnetic field using a current balance.
- Describe the concept of magnetic flux (Ø) as scalar product of magnetic field (B) and area (A) using the relation Ø = B.A.
- State and explain Ampere's law.
- Explain solenoid and toroid. Solve problems to obtain Magnetic flux density of solenoid and toroid by Ampere's law.
- Explain that a force acts on a charged particle in a uniform magnetic field.
- Solve problems using F = qvBsinθ.
- Solve problems using F = Ø = B.A.
- Describe a method to measure the e/m of an electron by applying magnetic field and electric field on a beam of electrons.
- Describe the motion of electrons in an electric field and magnetic field using a Cathode Ray tube.
- Solve problems using related equations.
- Derivation and use of τ = BANI.
- Describe the construction and working of Galvanometer and Its conversion into Voltmeter, Ammeter, and Avometer.
- Understand the turning effect on a current carrying coil in a magnetic field.
- Solve problems using: $R_x = (\frac{V}{I_g}) - R_g$ and $R_s = (\frac{I_g}{I - I_g})R_g$

Introduction:

A magnetic field is a fundamental concept in physics that describes the region around a magnet or a current-carrying conductor where magnetic forces are exerted on other magnets, conductors, or charged particles.

Permanent magnets, such as iron magnets, have their own intrinsic magnetic fields due to the alignment of their magnetic domains.

Magnetic fields have numerous practical applications, including:

- Electric motors
- Transformers
- MRI machines (Magnetic Resonance Imaging)
- Magnetic levitation trains
- Magnetic compasses
- Magnetic data storage (hard disk drives)

18.1 Magnetic Fields:

Magnetic field can be defined as ***"An area around a permanent magnet or a current carrying conductor, where they generate a magnetic force".*** Unlike in electrostatics, where electric charges create an electric field and exert electric forces on other charges, the magnetic field lacks a direct counterpart due to the absence of magnetic monopoles.

Magnetic fields can be generated through two primary methods:

- From permanent magnets.
- From current carrying conductors, also known as electromagnets

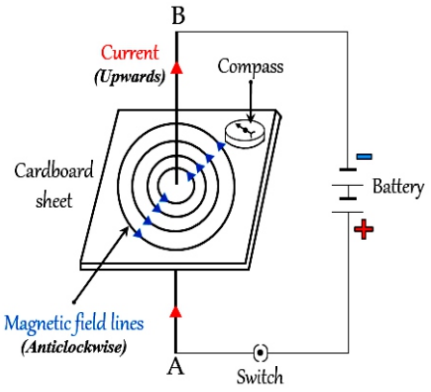


Figure 18. 1 (a) Current Carrying Wire

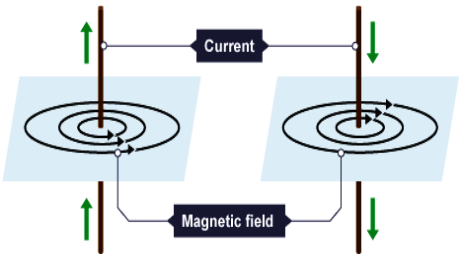


Figure 18.1 (b) Clockwise and Anticlockwise direction of magnetic fields

In 1819, Hans Christian Oersted made a significant discovery that revealed how a current Carrying conductor generates a magnetic field. The experiment depicted in Figure 18.1(a) clearly demonstrates this phenomenon. Simple setup consisting of a wire (AB) passing through a flat cardboard and connected to a battery via a switch. When the switch is closed, completing the circuit and enabling current flow through the wire, a circular magnetic field is generated around it. When iron fillings are scattered on to the cardboard, they become magnetized and arrange themselves in a circular pattern around the current-carrying wire. Additionally, a magnetic compass placed near wire AB indicates the direction of the magnetic field.

These observations illustrate that the magnetic field's direction produced by the current carrying wire follows the right-hand rule. According to this rule, when you grasp the wire with your right hand, your thumb points in the direction of the current, while the curl of your fingers indicates the direction of the magnetic field line. If the direction of the current is reversed, the needles also reverse their orientation. Conversely, if the current's direction is reversed, it generates a magnetic field in a clockwise direction. As shown in figure 18.1 (b).

18.2 Magnetic Force on a current carrying conductor:

Electric current is the orderly flow of charges through a conductor. As electrons move, they create a magnetic field around them, in addition to the electric field they produce. When it comes to a current-carrying conductor, the electric field generated by the moving electrons is counter balanced by the electric field produced by stationary protons within the conductor. Consequently, a magnetic field is produced around the current-carrying conductor.

When we place a current-carrying conductor with in a uniform External magnetic field, the interaction between the magnetic field produced by the conductor and the external magnetic field gives rise to external force, denoted as F, acting on the conductor, as illustrated in figure18.2.

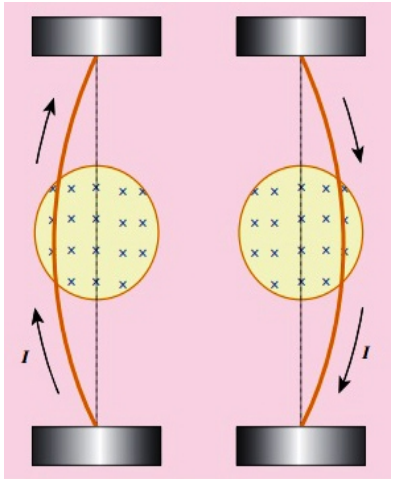


Figure 18.2 Magnetic Force on a current carrying conductor

A vertical wire is securely anchored at both ends and passes through a gap in a magnet. The external magnet generates a uniform magnetic field directed into the page. When a current flows upward through the wire, the wire deflects to the left. Conversely, reversing the current causes the wire to deflect to the right.

18.2.1 Factors on which the force acting on current carrying conductor in a magnetic field:

The force acting on a current-carrying conductor depends on several factors, such as, length, current, and the strength of the external magnetic field

The force (F) is directly proportional to the length of the conductor (L) that lies within the magnetic field

$$F \propto L$$

The force is also directly proportional to the current (I) passing through the conductor.

$$F \propto I$$

Similarly, the force is directly proportional to the strength of the external magnetic field (B).

$$F \propto B$$

Combining these factors, we will obtain the following relationship:

$$F \propto BIL$$

Here, 'k' is the proportionality constant, and in the SI unit system, its value is equal to 1.
Therefore, we can express the force as: $F = BIL$..... (18.1)
If the magnetic field is not perpendicular to the wire, Equation (18.1) can be modified to:

$F = I\ (L \times B)$ where 'X' denotes the vector cross product.

The magnitude of the force **(F)** is determined by: $F = BILSin\theta$
Where, θ represents the angle between the direction of the conductor's length (***L***) and the external magnetic field (***B***).

Maximum Force:

The deflecting force reaches its maximum when the angle (θ) between the length of the conductor and the external magnetic field is θ = 90 degrees, assuming other factors remain constant.

Minimum Force:

Conversely, the deflecting force is minimized when θ=0 degrees, signifying that the magnetic field and conductor are aligned.
Since magnetic force is a vector quantity, its direction can be determined using Fleming's left-hand rule, which provides a straight forward way to establish the direction of the force in relation to the current and magnetic field.

Fleming's left-hand rule:

Fleming's left-hand rule is used to determine the direction of force acting on a current carrying wire placed in a magnetic field, also to identify the direction of the forces in an electric motor in order to apply Fleming's left-hand rule, follow these steps: Extends our fore finger, thumb, and Middle finger such that they are mutually perpendicular as shown in figure 18.3.

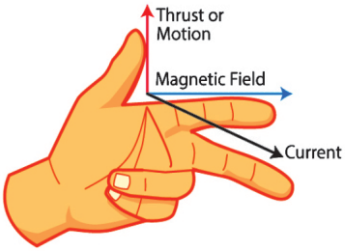


Figure 18. 3 Fleming's left-hand rule

Worked Example 18.1

A current carrying conductor of certain length placed at 90^0 to the magnetic field experiences a force F. What will be the force if the current is increased four times, length is halved and magnetic field is tripled?

Solution:

Step1: Write down the known quantities and quantities to be found.

I =4I, L = L/2, B= 3B $\acute{F}$=?

Step2: Write down the formula and rearrange if necessary.

As we know that force on current carrying wire placed in magnetic field is

$F=BIL\text{Sin}\theta$

When current is increased four times, length is halved and magnetic field tripled then,
Force,

$$\acute{F} = (4I)\ x\ \left(\frac{l}{2}\right)\ x\ (3B)\ F = 6\ F.$$

Result: Therefore, the force increases six times.

Worked Example 18.2

A conductor of length 50cm carrying a current of 5A is placed perpendicular to a magnetic field of induction 2×10^{-3} T. Find the force on the conductor.

Solution:

Step1: Write down the known quantities and quantities to be found.

I= *5A,* L = 50×10^{-2}m, B= 2×10^{-3}T, F =*?*

Step2: Write down the formula and rearrange if necessary.

Force on the conductor

$$F = BIL = 5 \times 50 \times 10^{-2} \times 2 \times 10^{-3}$$

$$F = BIL = 5 \times 10^{-3}\ N$$

Result: Therefore, F= 5×10^{-3} N

18.2.2 Magnetic Flux:

Magnetic flux is defined as **the number of magnetic lines of force passing through an area held perpendicular to it**. It is calculated by taking the dot product between the magnetic field and area vector. The dot product ensures that magnetic lines must pass perpendicular to the given area. Experimentally magnetic flux is usually measured with a Flux meter. Magnetic flux through an area A placed in uniform magnetic field (Figure18.4 is calculated by using formula as given below

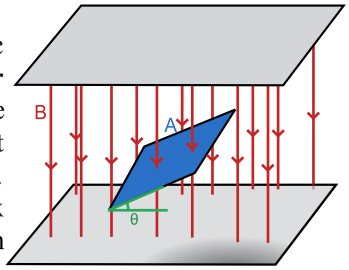


Figure 18.4 Magnetic Flux

$\emptyset_B = \mathrm{B}.\mathrm{A}$ **or** $\emptyset_B = \mathrm{BA}\cos\theta$ ...(18.2)

Where B is uniform magnetic field, **A** is an area vector whose magnitude is equal to the surface element and directed perpendicular to the given surface element.

Maximum Magnetic Flux:

The magnetic flux will be maximum when the angle between magnetic field **B** and area vector **A** is zero, i.e. $=0^0$ and surface will be perpendicular to the magnetic lines. This case is shown in figure 18.5(a).

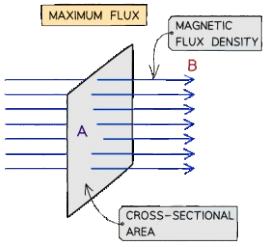


Figure 18.5 (a)
Maximum Magnetic Flux

Minimum Magnetic Flux:

Similarly when the angle between area vector and magnetic field is 90^0 then lines of force donot pass through the surface and magnetic flux will be minimum as shown in figure18.5 (b).

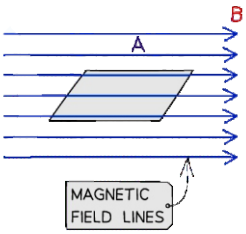


Figure 18.5 (b)
Minimum Magnetic Flux

Unit of Magnetic Flux:

The SI unit of magnetic flux is Weber (Wb) =

$$1Wb = \frac{Nm}{A}$$

It is named after renowned German Physicist and co-inventor of telegraph.

$\emptyset_B = \boldsymbol{B.A}$, represents the magnetic flux $\emptyset_B$ through a surface (**A**) in a magnetic field (**B**). To solve problems using this formula, you will typically need to work with scenarios where you know the magnetic field strength(B) and the area(A) through which the magnetic field lines pass.

Worked Example 18.3

Calculate the magnetic flux through a rectangular loop with dimensions 5 cm × 10 cm placed in a magnetic field of strength 0.2 T. The loop is oriented such that the magnetic field lines are perpendicular to the plane of the loop.

Solution:

Step1: Write down the known quantities and quantities to be found.

Magnetic field strength (B)=0.2T

$$Area\ of\ the\ loop\ (A) = 0.05\ m \times 0.10m = 0.005\ m^2$$

Step2: Write down the formula and rearrange if necessary.

$$\emptyset_B = BAcos\theta$$

Step 3: Put the values and calculate

$$\emptyset_B = BAcos\theta$$

$$\emptyset_B = \mathbf{0.2T \times 0.005m^2 = 0.001\ Wb\ (Weber)}$$

Result: So, the magnetic flux through the rectangular loop is 0.001 Weber.

DO YOU KNOW?

The magnetic properties of the disk's platters and the read/write head's ability to generate and detect magnetic fields play a crucial role in storing, encoding, and accessing data in computer systems.



18.2.3 Magnetic Flux Density:

Magnetic flux density is described in relation to the magnetic force experienced by a current carrying conductor placed perpendicularly to a magnetic field. For a uniform magnetic field, the flux density B is expressed by the equation: $B = \frac{F}{IL}$

Hence magnetic flux density can be defined as:

The magnetic flux density at a specific point in space is the force experienced per unit length by along straight conductor carrying unit current, placed perpendicular to the field at that particular location.

Magnetic flux density is most pronounced near the pole of a bar magnet and gradually decreases with distance from it.

Unit of Magnetic Flux Density:

The SI unit of magnetic flux density is Tesla (T), named after the Serbian-American inventor Nikola Tesla.

One Tesla: is defined as follows, If a conductor having length 1m and carries a current of 1 A placed perpendicularly to the magnetic field experience a force of one Newton then magnetic flux density will be 1Tesla.

In essence, the Tesla provides a standardized measure of the strength of a magnetic field, with larger values indicating stronger magnetic fields and smaller values representing weaker fields.

18.3 Measurement of Magnetic Flux density by current balance:

A simple experimental setup designed for the measurement of flux density between two magnets, As shown in figure 18.6. The magnetic field within this magnet configuration is uniform. To measure the length (L) of the current-carrying wire within this uniform magnetic field, a ruler can be used. When the wire carries no current, the magnet arrangement is positioned on top of balance, and the balance is calibrated to read zero. Subsequently, when an electric current (I) flows through the wire, the ammeter displays its magnitude. The wire experience an upward force, and in accordance with Newton's third law of motion, an equal and opposite force acts upon the magnets. Consequently, the magnets are pushed down wards, causing the balance to indicate a reading. The force (F) can be calculated as F = mg, with 'm' representing the mass indicated on the balance in kilograms and 'g' representing the acceleration due to gravity ($9.81 ms^{-2}$).

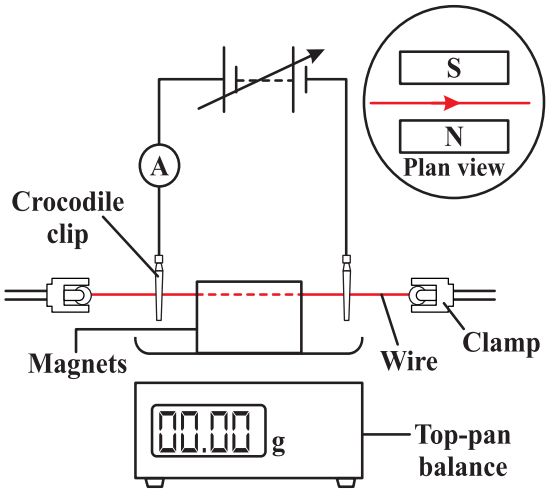


Figure 18.6
Experimental setup for the measurement of flux density

With the knowledge of F,I, and L, one can Determine the magnetic flux density (B) between the magnets using the following equation

$$B = \frac{F}{IL} \ldots\ldots\ldots (18.3)$$

18.4 Amperes Law:

Ampere's Law is an important principle in electromagnetism. It deals with the behavior of electric currents and their associated magnetic fields. It is named after Andre-Marie Ampere, a French physicist and mathematician who made significant contributions to the study of electromagnet is mid the 19th century.

It is known that an electric current flowing through a wire creates a magnetic field in its vicinity. When we envision a closed circular path with the wire positioned at its center, the Magnetic flux density (B) within this circular region changes, depending on both the current (I) and the distance (r) from the wire as shown in figure18.7.

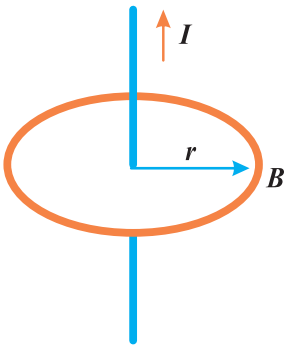


Figure 18.7
Current carrying wire

$$B \propto 2I \text{.....(i)}$$

$$\text{Or } B \propto \frac{1}{r} \text{.....(ii)}$$

Combing (i) and (ii), we get

$$B \propto \frac{2I}{r}$$

$$B = \frac{\mu_0 I}{2\pi r}$$

Where μ_0 is permeability of free space = $4\pi \times 10^{-7}$Wb A^{-1} m^{-1}

Now consider an Amperian loop in form of a closed path around the current carrying wire as shown in figure 18.7. For any path element ΔL we can write

$$B.\Delta L = \mu_0 I$$

Since angle between B and L is zero so $B.\Delta L = B\Delta L = \mu_0 I$

Now adding up over whole closed path

$$\sum B.\Delta L = \mu_0 I \ldots\ldots\ldots\ldots\ldots (18.4)$$

Equation (18.4) is the mathematical expression of Ampere law. Hence Ampere's law can be defined as the integral of the magnetic field (**B**) around a closed loop is directly proportional to the total electric current (**I**) passing through the loop.

Application of Ampere's Law:

(i) Solenoid:

A solenoid is along, tightly wound coil of wire, and often used to generate a uniform magnetic field within the coil's interior. It consists of many closely spaced turns of wire, and when an electric current flow through it, it creates a magnetic field that is similar to the field produced by a bar magnet. To determine the magnitude of B of a solenoid let us Consider an Amperian loop abcd with lengths $l_{ab}l_{bc}$, $l_{cd}$and l_{da}as shown in figure18.8.

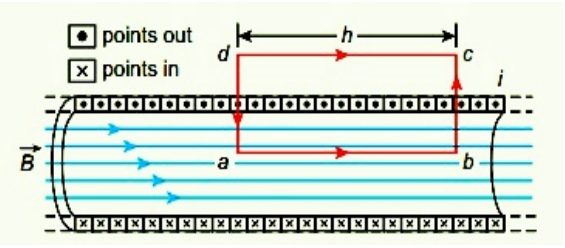


Figure 18.8 Amperian loop Solenoid

Then from Ampere's law

$$\sum B.\Delta l = \mu_o I$$

The left hand side of equation is

$$\sum B.\Delta l = B.l_{ab} + B.l_{bc} + B.l_{dc} + B.l_{da}$$

As the element al lengths along bc and da are perpendicular to the magnetic field so

$$B.l_{bc} = 0 and B.l_{da} = 0$$

Moreover, the magnetic field intensity outside of solenoid ≈ 0, so $B.l_{cd}=0$
Hence the magnetic field due to solenoid is given by

$$\sum B.\Delta l = B.l_{ab} = \mu_0 I$$

For N number of turns on the solenoid

$$\sum B.\Delta l = B.l_{ab} = \mu_0 NI$$

The number of turns per unit length is given by $\boldsymbol{n = \frac{N}{l_{ab}}}$

$$B = \mu_0 nI$$

Hence it is shown that B is independent of the position within the solenoid which shows that the field is uniform with in a solenoid.

Worked Example 18.4

Suppose we have a solenoid with 500 turns per meter and a current of 2.5 amperes flowing through it. Calculate the magnetic flux density inside the solenoid.

Solution:

Step1: Write down the known quantities and quantities to be found.

The no of turns = 500I = 2.5A

$\mu_0 = 4\pi \times 10^{-7}$

Magnetic flux density inside the solenoid,

(B) = ?

Step2: Write down the formula and rearrange if necessary.

$$\boldsymbol{B = \mu_0 nI}$$

Step 3: Put the values and calculate.

$\boldsymbol{B = \mu_0 nI}$

$B = 4\pi \times 10^{-7} \times 500 \times 2.5$

$B = 3.14 \times 10^{-3}$ T B = 3.14 mT

Result: So, the magnetic flux density inside the solenoid is 3.14 milli-tesla.

(ii) Toroid:

A toroid can be considered as a solenoid that is bent into the shape of a circle. It also generates a magnetic field within its interior. The magnetic field lines inside a toroid are generally circular and concentrated within the coil.

Suppose a toroid consist of N closely packed turns of copper wire and carry a current I as shown in 18.9.

Imagine a hollow circular ring wound with N number of turns of current-carrying wire. In this case, within the toroid, point P is our point of interest, and we will denote the magnetic field at this location as B. Inside the toroid, we construct an Amperian loop forming a circle that passes through point P. This

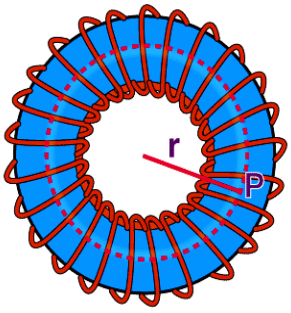


Figure 18.9 A toroid

arrangement gives rise to concentric circles within the toroid.
Because of the symmetrical nature of the magnetic field, the magnitude of the field at all points along the circle is uniform, and the field direction is tangential to the circle.
Therefore,

$$\sum B.\Delta l = \sum B\Delta l Cos(0) = B\sum \Delta l = (2\pi r)$$

Now, we can apply Ampere's law:

$$(2\pi r) = \mu_0 NI$$
$$B = \mu_0 \frac{NI}{2\pi r}$$

This is the expression for the magnetic field strength inside a toroid. It is inversely proportional to the radius r and directly proportional to the number of turns per unit length N and the current I flowing through the turns.

Worked Example 18.5

Suppose we have a toroid with 1000 turns and a current of 3 amperes flowing through it. The toroid has a circular cross-section with a radius of 0.1 meters. Calculate the magnetic flux density inside the toroid.

Solution:

Step 1: Write down the known quantities and quantities to befound.

The no of turns = 500I = 2.5A

$\mu_0 = 4\pi \times 10^{-7}$

Magnetic flux density inside the solenoid,

(B) = ?

Step 2: Write down the formula and rearrange if necessary.

$$B = \mu_0 \frac{NI}{2\pi r}$$

Step 3: Put the values and calculate

$$B = \mu_0 \frac{NI}{2\pi r}$$
$$B = (4\pi \times 10^{-7} \times 1000 \times 3)/(2\pi \times 0.1)$$
$$B = (4\pi \times 10^{-7} \times 1000 \times 3)T$$
$$B = 3.77 \times 10^{-3} T \text{ or } B = 3.77mT$$

Result: ***The magnetic flux density inside the toroid is 3.77milli-tesla.***

18.5 Force on a charged particle in a uniform magnetic field:

When a charged particle moves through a uniform magnetic field, it experiences a force known as the magnetic force. This phenomenon is described by the Lorentz force equation, which can be written as: $\vec{F}=(\vec{v}\times\vec{B})$.

In the equation, the magnitude of the charge (|q|): Determines the strength of the magnetic force. A higher charge magnitude results in a stronger force.

Velocity of the Particle (v):

The magnetic force depends on both the magnitude and direction of the velocity vector. If the particle is stationary (v = 0), it would not experience a magnetic force, as there is no velocity to interact with the magnetic field.

Magnetic Field (B):

The magnetic field vector, denoted as $\vec{B}$, represents the strength and direction of them agentic field in which the charged particle is located. A uniform magnetic field means that the strength and direction of the magnetic field are constant throughout the region. It is often represented by parallel magnetic field lines.

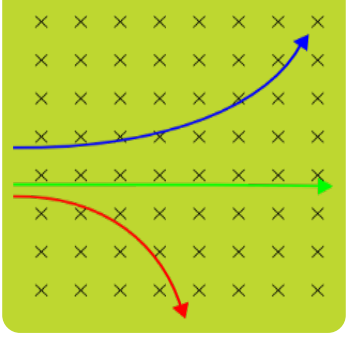
Cross Product ($\vec{v}\times\vec{B}$):

The key to understanding the magnetic force is the cross product $(\vec{v}\times\vec{B})$ in the Lorentz force equation. The cross product results in a force that acts perpendicular to both the velocity vector($\vec{v}$) and the magnetic field vector($\vec{B}$).The direction of this force is determined by the right-hand rule, which states that if you point your right thumb in the direction of the particle's velocity ($\vec{v}$) and your fingers in the direction of the magnetic field ($\vec{B}$) then the force ($\vec{F}$) will be directed perpendicular to both,either into or out of the palm of your hand, depending on the charge of the particle.

When a charged particle moves through a uniform magnetic field, a magnetic force is exerted on it. This force is always perpendicular to both the particle's velocity and the direction of the magnetic field. The charged particle will follow a curved path due to this force, known as a helical or circular trajectory, depending on the initial conditions and the specifics of the situation. The magnitude of the force depends on the charge, the velocity, and the strength of the magnetic field.

DO YOU KNOW?

When a proton, electron and neutron enters in uniform magnetic field points into the page then the neutron goes undeflected, the proton and electron moves upward and downward respectively



18.5.1 Charge to mass ratio of Electron:

Measuring the charge-to-mass ratio (e/m) of an electron is a classic experiment in the field of electromagnetism and particle physics. This experiment, known as the "e/m experiment, "involves applying both a magnetic field and an electric field to a beam of electrons. Here's a step-by-step method to measure the e/m of an electron:

Apparatus and Materials:

1. **Cathode Ray Tube (CRT)**: This is a vacuum tube that produces a beam of electrons. It consists of an electron gun, focus in, deflection plates, and a fluorescent screen as shown in figure 18.10.
2. **Magnetic Field Source**: We'll need a strong and uniform magnetic field source, such as a Helmholtz coil or a solenoid.
3. **Voltage Source**: A variable voltage source to create an electric field.
4. **Fluorescent Screen:** A screen coated with a phosphorescent material to visualize the electron beam.
5. **Rulers and Measurement Devices**: To measure the radius of the electron beam's circular path and the electric potential applied.

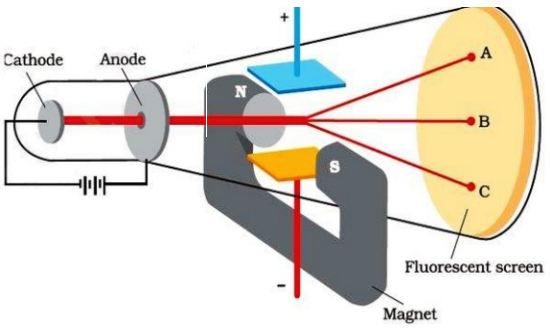


Figure 18.10 Cathode Ray Tube (CRT)

Experimental Procedure:

1. **Setup**:
   - Set up the CRT in a vacuum chamber to ensure the electron beam doesn't interact with air molecules.
   - Position the Helmholtz coil (or solenoid) around the CRT, providing a uniform magnetic field parallel to the beam path.
   - Connect the voltage source to the focusing and deflection plates to create an electric field perpendicular to the magnetic field.
2. **Calibration**:
   - Calibrate the magnetic field by measuring its strength at the location of the electron beam. We can do this using a magnetic field strength meter.
3. **Electron Beam Production**:
   - Apply a high voltage across the cathode and anode of the CRT to generate a beam of electrons from the cathode (electron gun).
   - Use the focusing and deflection plates to control and direct the electron beam.
4. **Observe the Electron Beam:**
   - Turn off the magnetic field and electric field to observe the electron beam's straight-line path on the fluorescent screen.
5. **Apply the Magnetic Field:**
   - Turn on the magnetic field, which causes the electron beam to bend in a circular path due to the Lorentz force (the interaction between the magnetic field and the moving electrons).
6. **Measure the Radius(r):**
   - Measure the radius of the circular path formed by the electron beam on the screen. Ensure the screen is marked with a scale for accurate measurement.

7. Apply the Electric Field:

- Apply a known electric potential (voltage) across the focusing and deflection plates perpendicular to the magnetic field. This will introduce an electric force on the electrons.

8. Adjust the Voltage:

- Adjust the voltage until the electron beam returns to its original, undeflected straight-line path.

9. Measure the Electric Potential (V):

- Measure the voltage applied across the plates.

Calculations:

An electron is released from cathode and it gains speed while passing through the potential difference *V*, hence gaining kinetic energy of 1eV. This electron kinetic energy is calculated by:

$$\frac{1}{2}mv^2 = V_e$$

The magnetic field produced by Helmholtz coils is perpendicular to this velocity, and produces a magnetic force which is transverse to both v and B: this provides the centripetal force makes an electron move along the circular trajectory; the radius of this trajectory r can be found from

$$\frac{mv^2}{r} = qvB$$

From this equation, we obtain the expression for the charge-to-mass ratio of the electron,

$$\frac{e}{m} = v/rB$$

Using the measured values for the radius (r), magnetic field strength (B), and electric potential (V), we can calculate the e/m ratio of the electrons which was found to be 1.7588×10^{11}C/kg

18.6 Measuring Instruments:

Galvanometer:

A Galvano meter is an instrument used to detect and measure electric currents. It is a highly sensitive electromagnetic apparatus capable of measuring even very small currents, such as those on the order of a few micro amperes.

Principle:

It works based on the principle of electromagnetic induction. When a coil carrying an electric current is positioned within an external magnetic field, it undergoes magnetic torque. The degree of deflection observed in the coil, caused by this magnetic torque, is directly proportional to the current's magnitude flowing through the coil.



Figure 18.11 Galvanometer

Construction:

1. **Coil**: The key component of a Galvano meter is a coil of wire (usually wound around as oft iron core) suspended within a magnetic field. The coil is mounted on a spindle so that it can rotate freely.
2. **Magnet**: A permanent magnet or an electromagnet is placed around the coil. The magnetic field lines from the magnet pass through the coil.
3. **Spring**: A delicate torsion spring is attached to the coil, providing a restoring torque when the coil is deflected.
4. **Pointer**: A thin pointer or needle is attached to the coil, allowing for the measurement of the deflection.

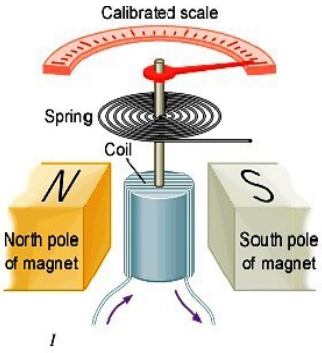


Figure 18.12 schematic diagram of Galvanometer

Working:

- When a small electric current flow through the coil, it generates a magnetic field around the coil according to Ampere's law. This magnetic field interacts with the external magnetic field (provided by the permanent magnet or electromagnet) to exert a torque on the coil.
- The torque causes the coil to rotate, and the amount of rotation is proportional to the current passing through it. This rotation is indicated by the deflection of the pointer on a calibrated scale.
- The coil continues to rotate until the restoring torque from the spring equals the torque due to the current-induced magnetic field. At this point, the pointer comes to rest, and its position on the scale indicates the magnitude of the current.
- Consider a rectangular coil consisting of N turns, each with a current I flowing through them and across-sectional area A. When this coil is situated within a uniform radial magnetic field B, it undergoes a torque as shown in figure 18.13.

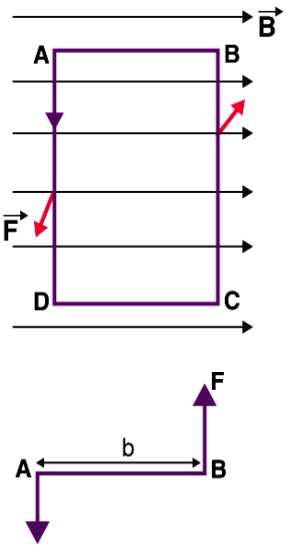


Figure 18.13 torque on rectangular coil

- Let's examine a single turn ABCD of the rectangular coil, characterized by a length L and breadth b as shown in figure 18.13. This turn is suspended within a magnetic field with strength of B, arranged so that the coil's plane is parallel to the magnetic field lines. As sides AB and DC are aligned parallel to the magnetic field, they donot experience any discernible force due to the magnetic field. However, sides AD and BC, which are perpendicular to the field's direction, encounter an effective force denoted as F, given by the equation F= BIl.
- By employing Fleming's left-hand rule, we can recognize that the forces acting on sides AD and BC are oriented in opposite directions to each other. When a pair of equal and

opposite forces, denoted as F and collectively referred to as a couple, act on the coil, they generate a torque. This torque induces a deflection in the coil.

- The torque(τ) is calculated as the product of the force(F) and the perpendicular distance (b) between these forces:$\tau = F \times b$
- Substituting the known value of F, we have:
  Torque(τ) acting on a single-loop ABCD of the $coil = BIL \times b$
  Where $L \times b$ represents the area (A) of the coil. Consequently, the torque acting on a coil with nturnsis given by:$\tau = NIAB$

This magnetic torque causes the coil to rotate, leading to the twisting of the phosphor bronze strip. Simultaneously, the spring (S) attached to the coil exerts a counter-torque, known as there storing torque (kθ), resulting in a stable angular deflection.

Under equilibrium conditions, we find:$k\theta = NIAB$

Here, k is termed the torsional constant of the spring, representing the restoring couple per unittwist. The deflection or twist (θ) is quantified as the reading displayed on a scale by a pointer connected to the suspension wire.

$$\theta = (NAB/k)I$$

Therefore $\theta \propto I$

The quantity (NAB/k) is a constant for a given galvanometer. Therefore, it is evident that the deflection observed in the galvanometer is directly proportional to the current flowing through it.

18.6.2 Conversion of Galvanometer to Ammeter:

To convert a galvanometer into an ammeter, a low-resistor known as a shunt is connected in parallel with the galvanometer. The selection of an appropriate shunt is based on the desired current range of the ammeter.

From the circuit as shown in figure 18.14.

- R_g represents the resistance of the galvanometer.
- G denotes the Galvanometer coil.
- I stand for the total current flowing through the circuit.
- I_g represents the total current passing through the galvanometer, which corresponds to the full-scale reading.
- R_S represent the value of the shunt resistance.

When the current Ig lows through theGalvano meter, the current passing through the shunt is determined by

$$I_s = I - I_g.$$

The voltages across both the galvanometer and shunt are equal due to their parallel connection.

Hence, we can establish the following equation:

$$R_g I_g = (I - I_g) \times R_s$$

$$R_s = \frac{R_g I_g}{I - I_g}$$



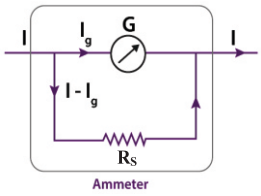


Figure 18. 14 Ammeter and its schematic diagram

Worked Example 18.6

A moving coil galvanometer of resistance 100Ω is used as an ammeter using a Shunt of 0.1Ω. The maximum deflection current in the galvanometer is 100μA. Find the current in the circuit so that the ammeter gives maximum deflection.

Solution:

Step1: Write down the known quantities and quantities to be found.

R_g=100Ω Rs= 0.1Ω, I_g=100μA

Step2: Write down the formula and rearrange if necessary.

$R_g \times I_g = (I - I_g) \times R_S$

Step 3: Put the values and calculate.

$$I = \frac{(R_g \times I_g + g \times Rs)}{Rs}$$

$$I = (1 + \frac{R_g}{Rs}) \times I_g$$

Result: We get $I = 100.1mA$

Conversion of Galvanometer to Voltmeter

A galvanometer is converted into a voltmeter by connecting it in series with high resistance as shown in figure 18.15. A suitable high resistance is chosen depending on the range of the voltmeter.



- ✓ R_g stands for the resistance of the galvanometer.
- ✓ R_X represents a high resistance component called series resistance.
- ✓ G denotes the galvanometer coil.
- ✓ I is the total current flowing through the circuit.
- ✓ I_g signifies the total current passing through the galvanometer, corresponding to a full scale deflection.
- ✓ V represents the voltage drop across the series connection of the galvanometer and the high resistance.

When the current I_g flows through the series combination of the galvanometer and the highresistance R, the voltage drop across the branch ab can be expressed as:

$$V = V_g + V_x$$

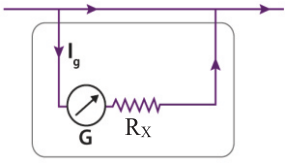


Figure 18. 15
Voltmeter and its schematic diagram

Current in series combination remain conserved therefore;

$$I_g = I_X$$

$$V = I_g R_g + I_g R_X$$

$R_x = V - I_g R_g / I_g$

$$R_x = \frac{V}{I_g} - R_g$$

This equation can be used to determine the value of RX.

Worked Example 18.7

A galvanometer coil of 40Ω resistance shows full range deflection for a current of 4mA. How can this galvanometer be converted into a voltmeter of range 0-12V?

Solution:

Step 1: Write down the known quantities and quantities to be found.

$R_g = 40\Omega$, V = range 0-12 V $I_g = 4\text{mA} = 4\times10^{-3}\text{A}$

Step 2: Write down the formula and rearrange if necessary.

$V = I_g(R_g+R_X)$

Step 3: Put the values and calculate.

$$R_X = \frac{V}{I_g - R_g}$$

$R_X = (12/(4\times10^{-3}))-40$

$R_X = 2960\Omega$

Result: we get $\mathbf{R_X = 2960\Omega}$

SUMMARY

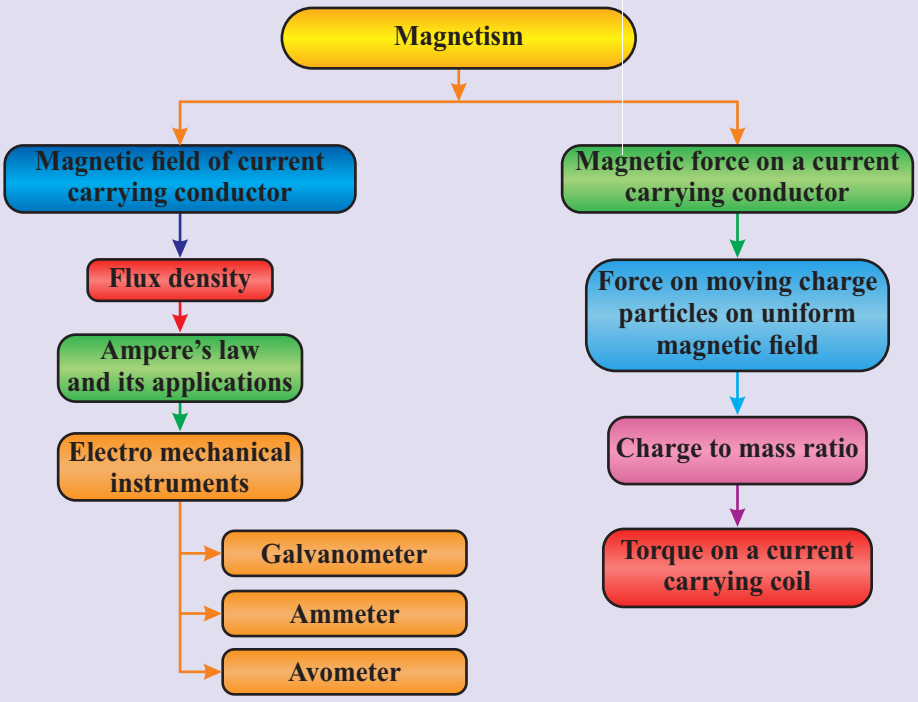
- ✓ Magnetic Field: The area around a permanent magnet or current-carrying conductor where a magnetic force is generated is called a magnetic field.
- ✓ Magnetic Force: The force exerted on a wire carrying current I in a uniform magnetic field is given by the formula: F=BILSinθ
- ✓ Measuring Magnetic Flux Density: A current balance can be used to measure magnetic flux density.
- ✓ Ampere's Law: This law quantifies how a current creates a magnetic field around it, providing a mathematical relationship between current and magnetic field strength.
- ✓ Solenoid: A long, tightly wound coil of wire used to generate a uniform magnetic field within the coil's interior. The magnetic field inside a solenoid is given by: $B=\mu_0 nI$
- ✓ Toroid: A solenoid curved to form a circular shape. The magnetic field strength inside a toroid is given by: $B = \frac{\mu oNI}{2\pi r}$ .This formula shows that the magnetic field strength is inversely related to the radius rand directly related to the number of turns per unit length N and the current I passing through the turns.
- ✓ The shunt resistance for a ammeter is given by

$$R_s = \frac{R_g I_g}{I - I_g}$$

- ✓ In case of voltmeter the high resistance is given by

$$R_X = V - \frac{I_g R_g}{I_g}$$

CONCEPT MAP



EXERCISE

Section (A): **Multiple Choice Questions (MCQs)**

Choose the correct answer:

1. If were verse the direction of electric current, then the direction of magnetic field will be:
   (a) Same (b) reversed
   (c) tangent (d) normal
2. The application of magnetic field is:
   (a) microwave oven (b) magnetic levitation trains
   (c) electrolysis (d) plant photosynthesis
3. The equation $F=BIL$ can only be used, if the magnetic field, length of conductor, and electric current are:
   (a) At right angles to each other (b) in same direction
   (c) Anti-parallel to each other (d) anti-perpendicular to each other

4. The charge on a particle is doubled, and its velocity remains the same, how then the magnetic force on the particle will be:
   (a) doubled (b) halved
   (c) Is the same. (d) quadrupled
5. Strength of magnetic field of solenoid can be increased by:
   (a) Increasing number of turns
   (b) decreasing number of turns
   (c) increasing the current through the solenoid
   (d) inserting a ferromagnetic core (e.g., iron) into the solenoid
6. The magnetic field inside along solenoid is:
   (a) equal to zero
   (b) uniform
   (c) decreases as we go away from the center to surface
   (d) increases as we go to wards the surface
7. The force between two current carrying conductor arises due to
   (a) Magnetic effect of current (b) Polarization
   (c) Electromagnetic induction (d) Electrostatic interaction
8. To measure a higher voltage, what should you do with the voltmeter's internal resistance?
   (a) Increase it (b) Decrease it
   (c) Keep it the same (d) It doesn't affect the measurement
9. A proton moves perpendicular to a uniform magnetic field. What is the direction of the force experienced by the proton?
   (a) Parallel to the magnetic field (b) In the direction of the proton's velocity
   (c) Perpendicular to the magnetic field and the proton's velocity
   (d) Opposite to the direction of the proton's velocity
10. An electron moves parallel to a uniform magnetic field. What is the magnitude of the force experienced by the electron?
    (a) Maximum, since the electron is moving in the same direction as the field
    (b) Minimum, since the electron is moving perpendicular to the field
    (c) Zero, since the electron is moving parallel to the field
    (d) Depends on the speed of the electron

Section (B): CRQs (Short Answered Questions):

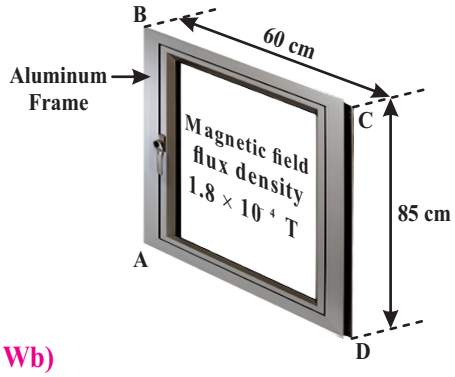
1. Charge particles are fired in vacuum tube hit a fluorescence screen. Will it be possible to know whether they positive or negative?
2. What is a solenoid, and how does it differ from a simple coil of wire?
3. Can a solenoid generate a magnetic field without any current flowing through it? Why or why not?
4. Explain why a toroid is often preferred over a straight solenoid when designing certain types of electrical components.
5. What role does the Ampere's circuital law play in understanding them agnetic field inside a toroid?
6. What is a galvanometer, and what is its primary function in an electrical circuit?

Section (C): ERQs (Long Answered Questions):

1. Can a galvanometer measure both DC and AC currents? Explain any limitations it might have with AC measurements.
2. Why should an ammeter ideally have a very low resistance compared to the circuit it is measuring?
3. What is the potential risk of using a voltmeter with a high internal resistance in a circuit? How can this risk be mitigated?
4. Describe the basic working principle of a voltmeter. How does it measure voltage across a component in a circuit?

Section (D): Numerical:

1. An aluminum window has a width of 60 cm and length of 85 cm as shown in the figure.
   a. When the window is closed the magnetic flux density is 1.8×10^{-4} Tisnormal to window.
   b. Calculate the magnetic flux through the window. **(9.18×10^{-5} Wb)**



2. The poles of a horse shoe magnet measures 8Cm × 3.2cm.the magnetic flux density between the magnet poles is 80mT. Outside of the magnet the magnetic flux density is zero. Calculate the magnetic flux density between the poles of a magnet.**(2.048×10^{-4} Wb)**
3. A wire 1.80 m long carries a current of 13.0 A and makes an angle of 35.0° with a uniform magnetic field of magnitude B = 1.50T. Calculate the magnetic force on the wire. **(19.46 N)**
4. A solenoid has length L = 1.23 m and inner diameter d = 3.55 cm, and it carries a current I= 5.57 A. It consists of five close-packed layers, each with 850 turns along length L. What is B at its center? **(Ans:24.2mT)**
5. A moving coil galvanometer has resistance of 50 Ω and it gives full scale deflection at4mA current. A voltmeter is made using this galvanometer and a 5KΩ resistance. Calculate the maximum voltage that can be measured using this voltmeter. **(Ans:20.2V)**
6. Compute the magnitude of the magnetic field of a long, straight wire carrying a current of 1A at distance of 1m from it. Compare it with Earth's magnetic field.
   (Ans:$2\times10^{-7}$and 100times than earth field)
7. Find the current in a long straight wire that would produce a magnetic field twice the strength of the Earth's at a distance of 5.0cm from the wire. (Magnetic field of Earth $=5.0\times10^{-5}$T). **(Ans:25A).**
8. What is flux density atadistanceof0.1m in air from along straight conductor carrying a current of 6.5 A. calculate the force per on a similar parallel conductor at a distance of 0.1m from the first and carrying a current of 3A. **(Ans: 13×10^{-6}Weber/$m^2$39$\times10^{-6}$N).**

Unit 19

Electromagnetic Induction

Teaching Periods 16 **Weightage %** 11



The Zorlu Energy Power Project is a wind farm situated in Jhimpir, Thatta District, in the Sindh province of Pakistan. This project, Pakistan's first wind farm, has gained global recognition. Wind power relies on electromagnetic induction to generate electricity.

In this unit student should be able to:

- Describe the production of electricity by magnetism.
- Explain that induced e.m.f can be generated in two ways.
  (i) by relative movement (the generator effect). (ii) by changing a magnetic field (the transformer effect).
- Infer the factors affecting the magnitude of the induced e.m.f.
- State Faraday's law of electromagnetic induction.
- Account for Lenz's law to predict the direction of an induced current and relate to the principle of conservation of energy.
- Apply Faraday's law of electromagnetic induction and Lenz's law to solve problems
- Explain the production of eddy currents and identify their magnetic and heating effects.
- Explain the need for laminated iron cores in electric motors, generators and transformers
- Define Self Induction and its unit
- How an inductor is used to store electric potential energy?
- Derive energy produced in Self Induction is $E = \frac{1}{2} LI^2$
- Explain Mutual Inductance (M) and its unit henry.
- Describe the construction of a transformer and explain how it works.
- Identify the relationship between the ratio of the number of turns in the primary and secondary coils and the ratio of primary to secondary voltages.
- Recall that how step up and step-down transformers can be used to ensure efficient transfer of electricity along cables.
- Describe the use of step-down and step-up transformers for the electric supply from power station to houses and electric appliances at home.
- Solve problems using $\frac{N_s}{N_p} = \frac{V_s}{V_p}$
- Define motional emf and Compute the potential difference across ends of a given rod or wire moving through a magnetic field
- Explain construction and working of an AC generator
- Identify the factors affecting induced EMF of an AC generator
- Solve problems using $\xi = \xi_0 \sin^2 \pi ft$
- Describe the main features of an A.C motor and the role of each feature.
- Explain the production of back emf in electric motors.

Introduction:

Electromagnetic induction stands as a cornerstone in the realm of physics, serving as a fundamental principle that explains the dynamic relationship between electricity and magnetism. Discovered by the famous physicist Michael Faraday in the early 19th century, electromagnetic induction encompasses the phenomenon where the change in the magnetic fields induces the generation of an electromotive force (EMF) or voltage within a conductor. This revolutionary concept not only laid the groundwork for the understanding of essential electrical processes but also paved the way for the development of modern technologies such as electric generators and transformers. As we examine the intricacies of electromagnetic induction, we work out the interplay between magnetic flux, electric current, and the connection that binds these phenomena together. In this chapter, we understand the implications and applications of electromagnetic induction, from its beginning to its current prevailing influence on modern technology and industry.

19.1.2 Production of electricity by magnetism:

In order to demonstrate that moving electric charges (i-e an electric current) give rise to magnetic force, put a magnet compass near to current carrying wire the needle of compass will be deflected as shown in figure 19.1 This shows that moving charges produce magnetic field and in turn deflect magnetic compass. Such a demonstration for the first time was carried out in 1819 by Professor Hans Oersted of the University of Copenhagen

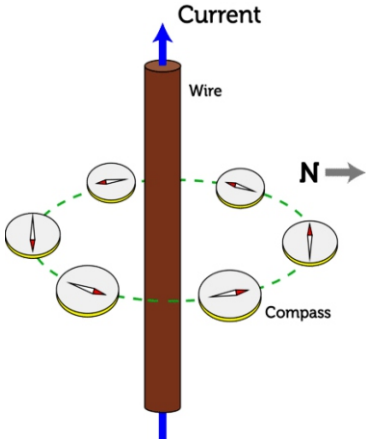


Figure 19.1 compass near to current carrying wire

The working of electric motor is based on this simple phenomenon, the current carrying coil produce magnetic field which interacts with the surrounding magnets causing the coil to rotate. In this case the electric energy is being converted into mechanical energy. The mechanical energy can be used to do variety of work, for example to pump water from ground level to a certain height.

On the other hand, the moving magnet produces electric force. To observe this phenomenon connect a coil with a sensitive galvanometer and a magnet move inside it. The relative motion of coil and motion will produce induced electric current. If the motion of magnet is stopped induced current ceases to exist.

The simple experiment as shown in figure 19.2 is similar to the working of electric generator. In the electric generator the conductor coil is rotated within the field of permanent magnet. The changing magnetic field produces induced electric current. In 1831 Micheal Faraday

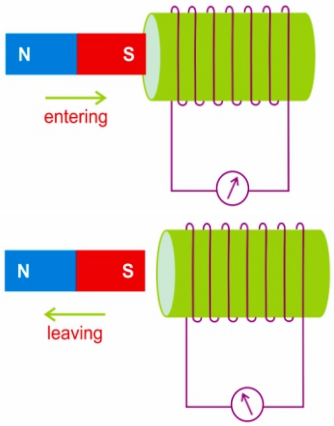


Figure 19.2 coil, Galvanometer and moving magnet

demonstrated that copper coils exposed to changing magnetic field produces electric current in the coil. Michael Faraday demonstrated this phenomenon with the help of copper coil, a magnet and a galvanometer. If the magnet was brought near to the coil an induced emf was produced and galvanometer shown current in one way. When the magnet was moved from the coil away the direction of induced current was reversed and galvanometer needle swing other way.

The phenomenon in which an emf is induced in the coil due to the change in linking magnetic flux is called electromagnetic induction.

DO YOU KNOW?

An induction stove, also known as an induction cooktop, is a type of cooking appliance that uses electromagnetic induction to heat cookware directly. Not all cookware is compatible with induction stoves. Cookware must have a magnetic bottom or base to work effectively on an induction cooktop.



19.1.1 Induced Electromotive force EMF:

Induced electromotive force (emf) can be generated in two primary ways: by relative movement and by changing a magnetic field.

(i) By Relative Movement (The Generator Effect):

This method is based on Faraday's Law of Electromagnetic Induction, which states that an emf is induced in a conductor when it experiences a change in magnetic flux. The generator effect, also known as motional emf, occurs when there is a relative motion between a conductor and a magnetic field.

Example:

In a simple electrical generator, a coil of wire rotates within a magnetic field. This induced emf drives an electric current through an external circuit.

(ii) By Changing a Magnetic Field (The Transformer Effect):

This method also relies on Faraday's Law but involves changing the strength or orientation of a magnetic field around a stationary conductor.

Example:

In a transformer, an alternating current in the primary coil creates a changing magnetic field. This is the basis for how transformers transfer electrical energy between circuits at different voltages.

Generator Effect:

Induced emf due to relative motion between a conductor and a magnetic field.

Transformer Effect:

Induced emf due to a changing magnetic field around a stationary conductor.

Both effects are fundamental principles in electromagnetism and are widely utilized in various electrical devices and systems.

19.1.4 Faraday's law of electromagnetic induction:

Faraday's law of electromagnetic induction is a fundamental principle in electromagnetism that describes the relationship between a changing magnetic field and the induced electromotive force (emf) or voltage in a closed circuit as illustrated in figure 19.3. It can be stated as follows:

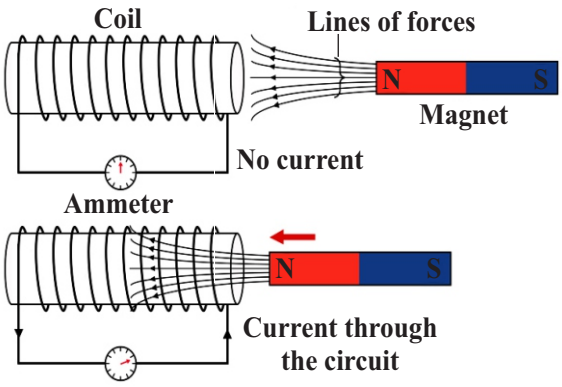


Figure 19.3 Electromagnetic induction

The electromotive force (emf) induced in a closed circuit is directly proportional to the rate of change of magnetic flux passing through the circuit.

Mathematically, Faraday's law is often expressed as:

$$\varepsilon = -N\frac{\Delta\emptyset}{\Delta t} \quad \ldots\ldots(19.\ 1)$$

Whereε represents the induced electromotive force (emf) measured in volts (V).

$\frac{\Delta\emptyset}{\Delta t}$ Represents the rate of change of magnetic flux with respect to time, it is measured in weber per second (Wb/s or V), and N indicates number of turns of the coil: Negative sign is introduced because the induced emf opposes the change in flux. It will be explained in Lenz's law.

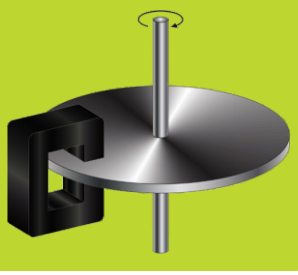
19.1.3 Factors affecting the magnitude of the induced emf:

The magnitude of the induced electromotive force (emf) in a circuit or coil is governed by several factors, which are primarily described by Faraday's law of electromagnetic induction. These factors include:

1. **Magnetic Flux Change:** If the rate of change of magnetic flux is higher, it will lead to a larger induced emf. Conversely, a slow change in the magnetic field will result in a smaller induced emf.
2. **Number of Turns in the Coil:** If a coil of wire has more turns, the greater will be the induced emf. Each turn of the coil contributes to the emf, so increasing the number of turns increases the overall emf.
3. **Area of the Coil:** The size of the coil or the area it encloses also affects the induced emf. A larger coil or a coil with a larger cross-sectional area will capture more magnetic flux lines, resulting in a larger induced emf.

DO YOU KNOW?

Electromagnetic braking is a method of slowing down or stopping a moving object, such as a vehicle, using the principles of electromagnetic induction. Electromagnetic braking is used in electric and hybrid vehicles, where regenerative braking systems utilize this technology to transform kinetic energy into electrical energy. This generated electrical energy is then stored in batteries or capacitors for future use.



4. **Angle between Magnetic Field and Coil:** The angle between the magnetic field lines and the plane of the coil affects the induced emf. When the magnetic field is perpendicular to the coil's plane (90°), the emf is maximized. When the field lines are parallel (0°), the emf is minimized.
5. **External Factors:** External factors, such as temperature, material property, pressure, and other environmental conditions, can also influence the induced emf, especially in situations where materials may exhibit nonlinear magnetic properties.

Self-Assessment Questions:

1. Explain Faraday's law of electromagnetic induction and its significance in electrical systems.
2. How does the rate of change of magnetic flux affect the induced electromotive force (EMF) in a coil?

19.1.4 Lenz's law and principle of conservation of energy:

After the introduction of Faraday's law of electromagnetic induction, Heinrich Friedrich Lenz formulated a rule for determining the direction of an induced current within a loop.

According to Lenz's law ***"The direction of induced emf in a circuit is always such that it opposes the cause which produces it"***

In Figure 19.4 (a) as the North Pole of the magnet approaches the surface of the coil, an induced current is generated within the coil in a anti clock wise direction. Consequently, the coil generates a North magnetic pole, resulting in a repulsive force experienced by the approaching magnet. In this manner, the coil resists the factor responsible for inducing the electromotive force (emf).

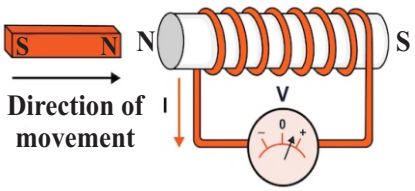


Figure 19.4 (a) Direction of induced emf

In Figure 19.3 (b), as the North Pole of the magnet moves away from the coil, an induced current is generated within the coil in a clockwise direction. Consequently, the coil generates a South magnetic pole, resulting in a attractive force experienced by the leaving magnet. Hence the law states that the induced current flows in a manner that oppose the change which is giving rise to it. That is why negative sign is introduced in equation 19.1.

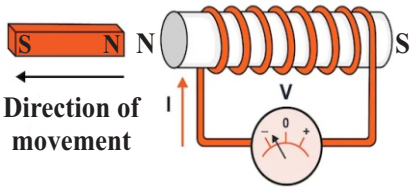


Figure 19.4 (b) Direction of induced emf

Lenz's law and Conservation of Energy:

The principle of conservation of energy states that energy cannot be created or destroyed; it can only change forms. When we apply this principle to electromagnetic induction, we see that the work done in changing the magnetic field is converted into electrical

energy in the form of the induced current. Hence lenz's law is in accordance with the law of conservation of energy.

Self-Assessment Questions:

1. Provide an example from everyday life or technology where Lenz's law is applied.
2. Discuss the consequences of not following Lenz's law in electromagnetic systems.

Worked Example 19.1

A coil with 100 turns and an area of 0.02 square meters is placed in a changing magnetic field. The magnetic field changes from 0.2 Tesla to 0.5 Tesla in 0.1 seconds. Calculate the induced electromotive force (emf) in the coil, the direction of the induced current, and how this relates to Lenz's law.

Solution:

Step 1: Write the known quantities and point out quantities to be found

B changes from 0.2 Tesla to 0.5 Tesla
Time t = 0.1 seconds.
No of turns = 100
Area A = 0.02 m^2
Induce emf = ?

Step 2: Write the formula and rearrange if necessary

$$\varepsilon = -N\frac{\Delta\emptyset}{\Delta t}$$

Step 3: Now, substitute the values into the formula:

$$\varepsilon = -N\frac{\emptyset_2 - \emptyset_1}{\Delta t} = 100 \times \frac{0.01 - 0.004}{0.1}$$

$$\varepsilon = -6\ V$$

Result: The induced emf in the coil is *6 V*, and the induced current flows counterclockwise to oppose the increase in the external magnetic field, in accordance with Lenz's law.

Worked Example 19.1

A coil with 200 turns and an area of 0.03 square meters is placed in a uniform magnetic field of 0.4 Tesla. The coil is quickly pulled out of the magnetic field in 0.2 seconds. Calculate the induced electromotive force (emf) in the coil and determine the direction of the induced current.

Solution:

Step 1: Write the known quantities and point out quantities to be found

B = 0.4T
Radius of coil = *r* =0.1m
$A = 0.03\ m^2$
The coil is pulled out of the magnetic field in 0.2 seconds.
Induce and its direction =?

Step 2: Write the formula and rearrange if necessary

Calculate the initial and final flux:

So initial flux = $\emptyset_1 = 0.4 \times 0.03 = 0.012$ Weber

Calculate the final flux:

When the coil is completely out of the magnetic field, the flux becomes zero because there is no magnetic field passing through the coil. Final flux $\emptyset_2 = 0$

Step 3: Now, calculate the rate of change of flux: =

$$\frac{\Delta\emptyset}{\Delta t} = \frac{\emptyset_2 - \emptyset_1}{\Delta t} = \frac{0 - 0.012}{0.2} = -0.06\ Wb/s$$

$$\varepsilon = -N\frac{\Delta\emptyset}{\Delta t} = -1.2\ Volts$$

Result: The induced current flows counterclockwise to oppose the decrease in magnetic flux, in accordance with Lenz's law.

Eddy currents and their magnetic and heating effects:

Eddy currents are circulating currents induced in a conductor when it is exposed to a changing magnetic field. They are a common phenomenon in electromagnetic systems. These currents can have both magnetic and heating effects. Now let's discuss about the production of eddy currents and their effects:

19.2.1 Production of Eddy Currents:

Eddy currents are produced according to Faraday's law of electromagnetic induction, which state that a changing magnetic field induces an electromotive force (emf) in a conductor. When a conductor, such as a metal plate or a coil of wire, is subjected to a changing magnetic field, the magnetic flux through the conductor changes, leading to the induction of eddy currents. As shown in figure 19.5.

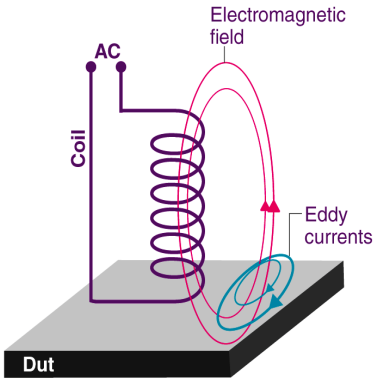


Figure 19.5
Production of Eddy Currents.

19.2.2 Magnetic Effects of Eddy Currents:

Counteracting Magnetic Field:

Eddy currents generate their own magnetic fields, and the direction of these fields opposes the original magnetic field that induced them as a result, eddy currents create a magnetic field that counteracts the original magnetic field's change, thereby reducing the net magnetic field.

Magnetic Damping:

In applications like electromagnetic brakes and magnetic dampers, eddy currents are intentionally induced to create a magnetic resistance that opposes motion. This magnetic damping effect is useful for controlling the movement of objects and slowing them down.

19.2.3 Heating Effects of Eddy Currents:

Joule Heating:

Eddy currents experience resistance as they flow through the conductor, and this resistance results in the conversion of electrical energy into heat, following Joule's law. Eddy currents can heat up the conductor. In some cases, such as induction heating for metal processing or cookware, this heating effect is used for practical applications.

19.2.4 Reduction of Eddy Currents:

In many electrical systems, eddy currents represent an undesirable source of energy loss, especially in transformers and electric motors. To minimize these losses, laminated cores are used in transformers to break up the conducting paths and reduce the formation of eddy currents.

The use of laminated iron cores in electric motors, generators, and transformers is essential for several important reasons:

In all three types of devices, the iron cores are exposed to alternating magnetic fields. When a solid iron core is used, it can create circular electrical currents within the core material called eddy currents. These eddy currents dissipate energy in the form of heat and can be highly inefficient. Laminated cores are constructed by stacking thin sheets or laminations of iron separated by insulating material. This design significantly reduces the formation of eddy currents because the laminations are electrically insulated from each other. Consequently, energy losses due to eddy currents are minimized.

Enhanced Efficiency:

By reducing the losses associated with eddy currents, laminated cores improve the overall efficiency of electric motors, generators, and transformers. Less energy is wasted as heat, allowing these devices to operate more efficiently and with less energy consumption.

Mitigation of Vibration and Noise:

Eddy currents generated in a solid iron core can lead to vibrations and noise, which can be undesirable in many applications. Laminated cores help reduce these vibrations and noise levels, making the devices quieter and more mechanically stable.

Better Cooling and Thermal Management:

Since laminated cores result in reduced heat generation due to decreased eddy current losses, they often allow for more efficient cooling and thermal management in these devices. This can lead to longer operational lifetimes and improved reliability.

Self-Assessment Questions:

1. When might the heating effects of eddy currents be undesirable?
2. How do the dimensions of a conductor impact the formation and magnitude of eddy currents?

19.3.1 Self Induction:

When an electric current pass through a coil, it creates a magnetic field around it. If the current in the coil changes, either increasing or decreasing, the magnetic flux within the coil also changes accordingly. As a result of the change in magnetic flux, an induced emf is generated in the same coil. This process, where a changing current in the coil induces an emf in the coil itself, is known as self-induction.

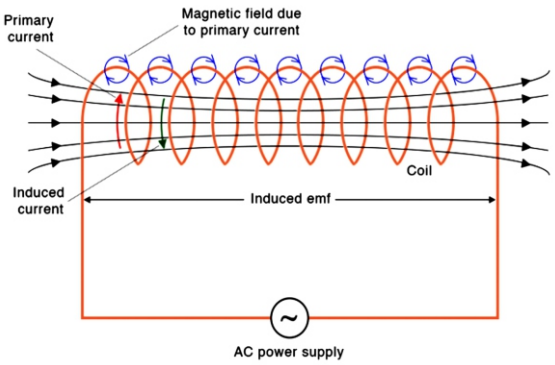


Figure 19.6 Self Inductions

Consider the circuit shown in Figure 19.6. which comprises a coil connected in series with a battery and a rheostat. When the rheostat is adjusted, it changes the current flowing through the circuit, leading to changes in the magnetic flux within the coil. These variations in magnetic flux result in the induction of an electromotive force (emf) within the coil.

If we denote the magnetic flux through a single loop of the coil as $\emptyset$, then the total magnetic flux passing through the N turns of the coil would be $N\emptyset$. This relationship holds true because the magnetic flux, directly proportional to the magnetic field, which, in turn, is proportionate to the current (I) flowing through the circuit

$$N\emptyset \propto I$$

$$N\emptyset = LI \ldots\ldots\ldots\ldots \text{(19.2)}$$

Where $L = \frac{N\emptyset}{I}$ is constant of proportionality called the self inductance of coil. It depends on factors such as the coil's number of turns, its cross-sectional area, and the material used for the core. Hence inductance of coil made up of soft iron is greater than the air core coil.

By the Faradays law of electromagnetic induction

$$\varepsilon = -N\frac{\Delta\emptyset}{\Delta t} \ldots\ldots\ldots\ldots\ldots\ldots \text{(19.3)}$$

Substitute equation (19.2) in Equation (19.3), we get

$$\varepsilon = -L\frac{\Delta I}{\Delta t} \ldots\ldots\ldots\ldots\ldots \text{(19.4)}$$

Equation (19.4) shows that the self-induced electromotive force (emf) within a coil is directly proportional to the rate of change of current within the coil. We can define the self-inductance, denoted as "L," as the ratio of this emf to the rate of change of current in the coil.

The negative sign in equation (19.4) signifies that the self-induced emf always opposes the cause that produces it. This is the reason why the self-induced emf is often referred to as the "back emf." Consequently, due to the phenomenon of self-inductance, coils of wire are commonly known as "inductors." In the field of electrical and electronics engineering, inductors find widespread use.

Unit of Inductance:

The unit of self-inductance is the Henry (H), named after the American physicist Joseph Henry.

One Henry (1 H) is defined as the amount of self-inductance in a circuit when a change in current of 1 ampere per second (1 A/s) induces an electromotive force (emf) of 1 volt (1 V) within the same circuit. Mathematically, this relationship can be expressed as:

$$1\ Henry = \frac{1 Volt}{1\ Ampere\ per\ Second}$$

19.3.2 Energy Stored in an Inductor:

An inductor is a passive electrical component that stores energy in the form of a magnetic field when an electric current flow through it. The ability of an inductor to store electric potential energy is based on the fundamental principle of electromagnetic induction. Here's a simple explanation of how an inductor stores electric potential energy:

Current Flow:

When an electric current flow through a coil of wire (the inductor), it creates a magnetic field around the coil.

Magnetic Field Buildup:

As the current increases, the strength of the magnetic field around the coil also increases. This process takes a short amount of time, as the magnetic field does not build up instantly; it grows with the rate of change of the current.

Energy Storage:

The energy is stored in the magnetic field. The inductor stores electric potential energy in this magnetic field. The more current that flows through the inductor or the faster the current changes, the stronger the magnetic field, and thus, the more energy is stored.

Magnetic Field Collapse:

When the current decreases or stops (like when you turn off the power), the magnetic field starts to collapse. As it collapses, it induces an electromotive force (EMF) or voltage in the coil.

Released Energy:

This induced voltage represents the stored energy being released. The inductor converts the stored magnetic energy back into electric energy. This energy can be used to sustain the current for a short period or can be transferred to other parts of the circuit.



Self-Assessment Questions:

1. When is the energy stored in an inductor released?
2. Provide examples of practical applications where the energy stored in inductors is crucial.

19.3.3 Energy produced in Self Induction:

An inductor stores energy in its magnetic field when it carries a direct current (DC). This energy remains stored as long as the inductor continues to carry the current. When the current in the inductor increases, the stored energy also increases, and conversely, it decreases when the current is reduced.

Consider an inductor connected to a DC power source through a switch, as depicted in Figure 19.7. When the switch is closed, the current in the inductor gradually rises until it reaches its maximum value, denoted as I. This changing current leads to a corresponding change in the magnetic flux within the coil, causing an induced electromotive force (emf) to establish itself in the coil. Consequently, an induced current is generated in the circuit, which works to minimize the current produced by the battery in accordance with Lenz's law.

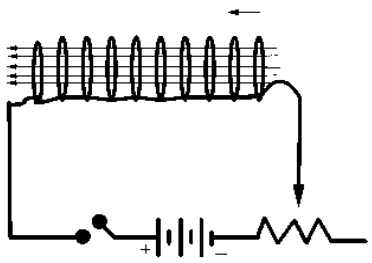


Figure 19.7 Energy produced in Self Induction

As a result, the battery must perform work on the charges to build up the current I. This work is mathematically expressed as follows:

$$W = \Delta V \Delta q \quad \ldots\ldots\ldots\ldots (19.5)$$

As the induced emf is produced in the inductor so $\Delta V = \mathcal{E}_L = L\frac{\Delta I}{\Delta t}$ and using work energy relation equation (19.5) can be written as

$$E = \frac{\Delta q}{\Delta t} L \Delta I \quad \ldots\ldots\ldots\ldots (19.6)$$

Since $\frac{\Delta q}{\Delta t} = \frac{0+I}{2} = \frac{I}{2}$

Equation (19.6) becomes $E = \frac{1}{2} L I^2$

Where:

- E is the energy stored in joules (J).
- L is the self-inductance of the inductor in Henries (H).
- I is the current flowing through the inductor in amperes (A).

This formula tells you that the energy stored in an inductor is directly proportional to the square of the current passing through it and is also dependent on the inductance of the inductor itself. As the current increases, the energy stored in the inductor increases, and as the current decreases, the stored energy decreases.

19.4.1 Mutual Inductance:

Imagine two closely positioned coils, as depicted in Figure 19.8. The first coil, known as the primary coil, is connected to a battery through a variable resistor 'R' and a switch 'S.' The second coil, known as the secondary coil, is connected to a galvanometer.

When the resistance is adjusted, it causes a change in the

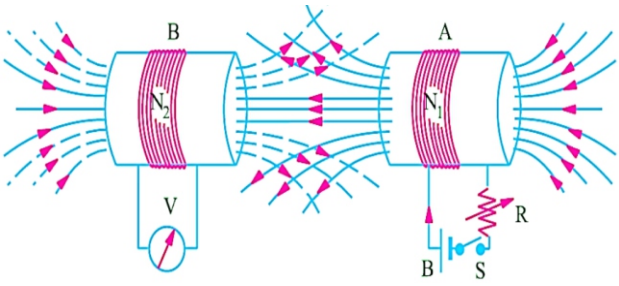


Figure 19.8 Mutual Inductance

electric current 'I' within the primary coil. This change results in the production of a varying magnetic field that interacts with the nearby secondary coil, effectively establishing a connection between them. Consequently, an electromotive force (EMF) is induced in the secondary coil in accordance with Faraday's law. **"A phenomenon where a changing current in one coil induces an electromotive force (EMF) in another coil is known as mutual induction".**

Consider the two closely wound coils of wire shown in Figure 19.8. The current I_1 in primary coil A, which has N_1 turns, creates a magnetic field. Some of the magnetic field lines pass through secondary coil B, which has N_2 turns. Let the flux passing through secondary coil having N turns be is $N_s\emptyset_s$. This magnetic flux depends on the primary coil current I_p

$$N_s\emptyset_s \propto I_p$$

$$N_s\emptyset_s = MI_p \ldots\ldots\ldots\ldots (19.7)$$

$$M = \frac{N_s\emptyset_s}{I_p}$$

Here, 'M' represents the proportionality constant known as the mutual inductance between the two coils. Mutual inductance depends on factors such as the number of turns in the coils, their geometrical characteristics, the materials used to construct the coils, and the separation between them. When the separation between the coils increases, the mutual inductance decreases as a result of the diminishing magnetic flux.

From faradays law of electromagnetic induction

$$\varepsilon_s = -N_s\frac{\Delta\emptyset_s}{\Delta t} \ldots\ldots\ldots (19.8)$$

Substitute Equation (19.7) in Equation (19.8), we get

$$\varepsilon_s = -M\frac{I_p}{\Delta t} \ldots\ldots\ldots\ldots (19.9)$$

Equation (19.9) shows that induced emf produce in the secondary coil is equal to negative rate of change of current in the primary coil.

Equation (19.9) can be used to provide us a definition of mutual inductance which is Henry

$$M = -\frac{\varepsilon_s}{\left(\frac{I_p}{\Delta t}\right)}$$

One Henry (H) of mutual inductance can be defined as the mutual inductance between two coils when an EMF of one volt is induced in the secondary coil when the rate of change of current in the primary coil is one ampere per second.

Mutual inductance measures the ability of one coil to induce an electromotive force (EMF) in another coil when the current in the first coil changes.



Self-Assessment Questions:

1.Give an example of devices or systems where mutual induction is utilized.

2. What are the advantages of using mutual induction in practical applications.

19.5.1 Transformer:

A transformer is an electrical device used to transfer electrical energy between two or more coils of wire through electromagnetic induction. It can either step up (increase) or step down (decrease) the voltage of an alternating current while keeping the frequency of the alternating current unchanged. It works on the principle of mutual induction.

Construction of a Transformer:

The transformer consists of following main components as shown in figure 19.9.

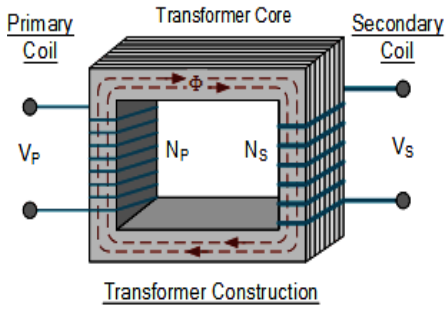


Figure 19.9 Transformer:

Core:

Transformers consist of two coils of wire, known as the primary coil and the secondary coil, wound around a common magnetic core. The core is typically made of materials with high magnetic permeability, such as laminated iron or ferrite, to enhance the efficiency of the transformer.

Primary Coil:

The primary coil is connected to the input voltage V_p source. It consists of a specific number of turns of wire wound around one section of the core denoted as N_p.

Secondary Coil:

The secondary coil is connected to the load or the device that needs the transformed voltage. It has a different number of turns of wire wound around another section of the core denoted by N_s.

Insulation:

Both the primary and secondary coils are insulated from each other to prevent electrical contact and ensure electrical isolation.

Working of a Transformer:

Imagine we have an alternating electromotive force (emf) applied to the primary coil. If, at a certain moment (t), the magnetic flux within the primary coil is changing at a rate of $\frac{\Delta\emptyset}{\Delta t}$ this change in flux will induce a counteracting electromotive force (emf) in the primary coil, opposing the applied voltage. We can express the instantaneous value of this self-induced emf as follows:

$$\text{Self-induced emf} = -N_p \frac{\Delta\emptyset}{\Delta t}$$

If the coil's resistance can be considered negligible, then the opposing emf in the primary coil is equal in magnitude but opposite in direction to the applied voltage. This relationship can be represented as:

$$V_p = -N_p \frac{\Delta\emptyset}{\Delta t} \ldots\ldots\ldots (19.10)$$

Here, N_p represents the number of turns in the primary coil.

Now, let's assume that the flux passing through the primary coil also flows through the secondary coil. In other words, these two coils are magnetically linked. Hence, the rate of change of magnetic flux in the secondary coil will also be $\frac{\Delta\emptyset}{\Delta t}$, and the magnitude of the induced emf across the secondary coil can be expressed as:

$$V_s = -N_s \frac{\Delta\emptyset}{\Delta t} \text{ ..... (19.11)}$$

Where *Ns* is number of turns in secondary

When we divide equation (19.11) by equation (19.10), we obtain

$$\frac{V_s}{V_p} = \frac{N_s}{N_p}$$

Step-Up or Step-Down Transformers:

The ratio of the number of turns in the primary coil (Np) to the number of turns in the secondary coil (N_s) determines whether the transformer is a step-up or step-down transformer. If $N_s > N_p$ than it's a step-up transformer and it increases the voltage. Conversely, if $N_p > N_s$ it is a step-down transformer and it decreases the voltage.

19.5.2 Step up transformer:

Step-up transformers are designed to increase the voltage of electricity. They have more turns in the secondary coil than in the primary coil, resulting in a higher secondary voltage compared to the primary voltage.

Reduced Current:

Increasing the voltage through a step-up transformer reduces the current while maintaining the same power (P = VI). Lower current reduces the resistive losses in the transmission cables. The power lost as heat (I^2R losses) is proportional to the square of the current, so decreasing the current significantly reduces these losses.

Efficient Long-Distance Transmission:

High voltage is essential for efficient long-distance power transmission. Step-up transformers are used at power generation plants to raise the voltage, allowing electricity to be transported over extensive electrical grids with minimal energy loss. This is especially important for transmission lines covering large distances.

19.5.3 Step-Down Transformers:

Voltage Reduction:

Step-down transformers, on the other hand, lower the voltage from high levels to safer, more usable levels for homes and industries. They have fewer turns in the secondary coil, which decreases the secondary voltage.

Balanced Current:

While step-down transformers decrease voltage, they increase the current to maintain the same power. This is beneficial for local distribution because it ensures that the power can be delivered to homes, businesses, and industrial facilities while minimizing I^2R losses over shorter distances.

Self-Assessment Questions:

1. Differentiate between a step-up transformer and a step-down transformer, providing examples of each.
2. Name the fundamental principle behind the operation of a transformer

19.5.4 Transmission of Electricity:

Step-down and step-up transformers are used for the electric supply from power station to houses and electric appliances. Here's how they are used in the electric supply process a shown in figure 19.10:

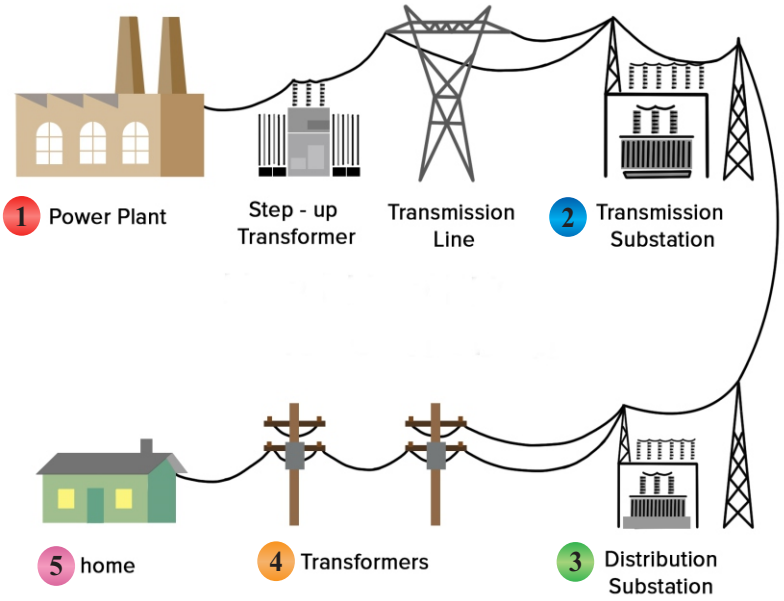


Figure 19.10 Transmission of Electricity

1. Power Generation at the Station:

Electricity is generated at power stations, often using generators powered by various sources such as hydel, coal, natural gas, nuclear energy, or renewable sources like wind and solar. The electricity generated is typically produced at a relatively low voltage level. For efficient generation and long-distance transmission, higher voltages are preferred to minimize energy losses.

2. Step-Up Transformers at the Power Station:

To raise the voltage to levels suitable for long-distance transmission, step-up transformers are employed at the power station. Step-up transformers increase the voltage and reduce the current, which lowers the I^2R losses, making electricity transmission more efficient. High-voltage transmission lines, such as overhead power lines or underground cables, carry the electricity over long distances to substations.

3. Transmission and Substations:

The high-voltage electricity is transmitted over transmission lines to substations located at various points in the electricity distribution network. At these substations, step-down transformers are used to lower the voltage to a more manageable level for further distribution.

4. Step-Down Transformers in Substations:

Step-down transformers reduce the high-voltage electricity to lower, safer voltage levels, suitable for local distribution. This lower voltage is then distributed to houses through distribution lines.

Worked Example 19.3

Suppose a step-up transformer have 200 turns in the primary coil and 800 turns in the secondary coil. If the voltage in the primary coils is 120 volts, calculate the voltage in the secondary coil.

Solution:

Step 1: Write down the known quantities and quantities to be found.

$N_p = 200$ $N_s = 800$ $Vp = 120$ volts $Vs = ?$

Step 2: Write down the formula and rearrange if necessary.

Using the turns ratio equation: $\frac{V_s}{V_p} = \frac{N_s}{N_p}$

Step 3: Substitute the known values:

Substitute the known values: $\frac{V_s}{120} = \frac{800}{200}$

Step 4: Solve for V_s,

We get: $V_s = 480\ Volts$

Result: The voltage in the secondary coil is 480 volts.

Worked Example 19.4

Let a step-down transformer have 600 turns in the primary coil and 150 turns in the secondary coil. The voltage in the primary coil is 240 volts. Calculate the voltage in the secondary coil.

Solution:

Step 1: Write down the known quantities and quantities to be found.

$N_p = 600$ $N_s = 150$ $Vp = 240$ $Vs = ?$

Step 2: Write down the formula and rearrange if necessary.

$\frac{V_s}{V_p} = \frac{N_s}{N_p}$

Step 3: Substitute the known values:

$$\frac{V_s}{240} = \frac{150}{600}$$

Step 4: Now, solve for V_s, we get:

$V_s = 60$ Volts

Result: So, the voltage in the secondary coil is 60 volts.

19.6.1 Motional emf:

Motional electromotive force (emf) is a phenomenon that arises when a conductor moves through a magnetic field, inducing an electromotive force within the conductor. This concept is based on Faraday's law of electromagnetic induction and is a fundamental aspect of electromagnetism. Motional emf is commonly appears in various practical applications. For example, the generation of electricity by a generator. When a coil of wire rotates in a magnetic field, motional emf is induced, leading to the generation of electric current.

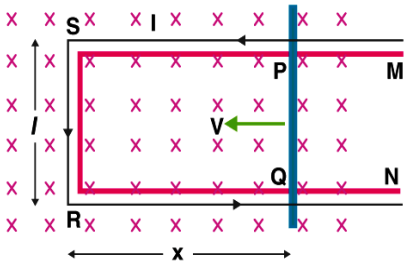


Figure 19.11 Motional emf

Understanding motional emf is crucial in designing and analyzing systems involving the motion of conductors in magnetic fields, and it plays an important role in the working of devices such as generators and certain types of sensors.

In summary, the Motional electromotive force (emf) is the voltage induced in a conductive rod as it moves through a magnetic field.

Consider a straight wire PQ having length l, as shown in figure 19.11. The conductor is in motion within a rectangular loop PQRS. This motion occurs within a uniform magnetic field B, perpendicular to the plane of the page. We assume that the rod moves uniformly at a constant velocity of v meters per second, and the surface it moves on is frictionless.

Each free electron of the conductor is moving with the conductor and thus experiences a force

$$F = -e(\vec{V} \times \vec{B}) \quad \textbf{or} \quad F = e(\vec{V} \times \vec{B})$$

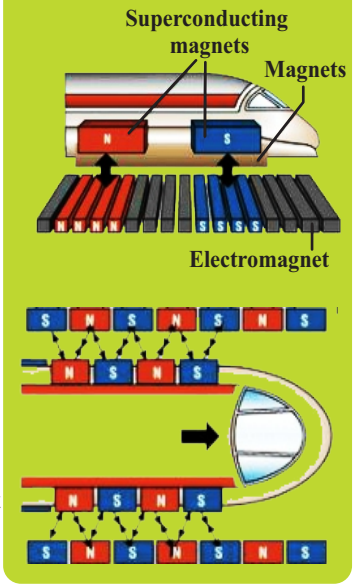
Electrons continue to accumulate at one end of the conductor, creating an excess of electrons, which results in a negative charge at that end. Conversely, the opposite end, with a deficiency of electrons, leads to the development of a positive charge. This process generates a potential difference within the wire, which is termed as motional electromotive force (emf). In order to derive an expression for the motional emf, we use the definition of potential difference.

Since the potential difference is defined as work done per unit charge

$$V = \frac{w}{q} = \frac{Fl}{q} = \frac{qvB\sin\theta \times l}{q} = vBl\sin\theta$$

DO YOU KNOW?

Magnetic Levitation (Maglev) fastest trains in the world use the repelling force of magnets to stay in the air. So, they do not even have wheels and also do not touch the track. Maglev trains use superconducting magnets and motional emf to levitate and propel the train along the tracks.



If the conductor is moving right angle to the magnetic field then $\theta = 90^0$

$$V = vBl \ldots\ldots\ldots 19.12$$

Where **V** is motional emf, v represents the velocity of conductor moving in a magnetic field and $\boldsymbol{l}$ is the length of conductor

Self-Assessment Questions:

1. Describe Fleming's right-hand rule and its application in the context of motional EMF.
2. How does the direction of motion, magnetic field, and current relate in Fleming's right-hand rule?

Worked Example 19.5

A straight conductor of length L = 0.5 meter move at a velocity of 2 m/s perpendicular to a magnetic field B = 0.1 Tesla. Calculate the motional emf induced in the conductor.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Length of the conductor, L = 0.5 meter

Velocity of the conductor, v =2 m/s

Magnetic field strength, B = 0.1 Tesla

Step 2: Write down the formula and rearrange if necessary.

$V = vBl$

Step 3: Substitute the known values:

V = (0.1 T) · (0.5 m)·(2 m/s)

V = 0.1V

Step 4: Therefore, the motional emf induced in the conductor is 0.1 volts.

Result: This means that as the conductor moves through the magnetic field, an electromotive force of 0.10 volts is induced across its ends, creating the potential for an electric current if the circuit is closed.

19.6.2 AC Generator

An AC generator, also known as an alternator, is a device that converts mechanical energy into electrical energy in the form of alternating current (AC). It's based on the principle of electromagnetic induction, which was first described by Michael Faraday.

Construction of an AC Generator:

An AC generator typically consists of the following four components as shown in figure 19.12:

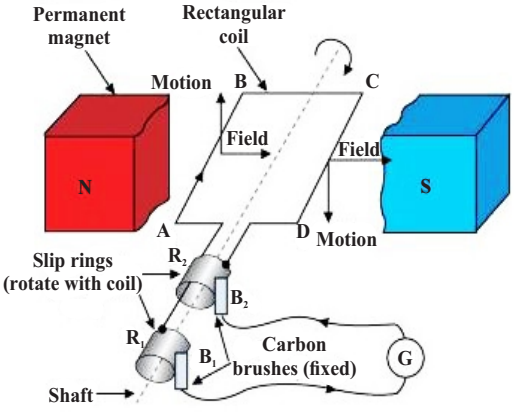


Figure 19.12 AC generator

(i) Armature
(ii) Field magnet
(iii) Slip-rings
(iv) Brushes

Armature:

The armature or rotor is a rectangular coil mounted on a rotating shaft. It is the component that spins within a magnetic field to induce electrical current.

Permanent magnet:

The permanent magnet or stator is a stationary part of the generator that produces a magnetic field. It is typically made up of a set of magnets or electromagnets arranged in a circular or cylindrical configuration around the rotor.

Slip Rings and Brushes:

The ends of the rotor coil are connected to slip rings, which are conductive rings that rotate with the rotor. Brushes, typically made of graphite, press against the slip rings to collect the generated electrical current.

Shaft and Bearings:

The rotor is mounted on a shaft that allows it to rotate freely. Bearings are used to reduce friction and enable smooth rotation.

Working of an AC Generator:

The working of an AC generator involves several steps:

Rotation of the Armature coil:

The coil or rotor is mechanically rotated using an external source of power, such as an engine, a turbine, or any other energy source.

Generation of a Changing Magnetic Field:

As the rotor spins within the stator's magnetic field, the magnetic field within the rotor coil changes. This changing magnetic field induces an electromotive force (emf) or voltage in the coil,

Generation of Alternating Current:

The induced voltage causes an electric current to flow through the coil, and since the magnetic field's direction is changing, the current produced is alternating in nature. This means the direction of the current constantly switches back and forth, resulting in an AC output.

Collection of Output:

The alternating current generated in the rotor coil is collected using slip rings and brushes. The brushes maintain contact with the slip rings as they rotate, allowing the generated AC to be drawn from the generator.

Induced EMF of an AC generator:

Suppose the armature coil AHCD rotates counterclockwise. As it rotates, the magnetic flux linked with it changes, inducing a current in the coil as shown in figure 19.13(a,b). The direction of the induced current follows Fleming's right-hand rule.

When the armature is in the vertical position and rotates counterclockwise, wire AH moves downward, while DC moves upward. This causes the induced electromotive force (emf) to flow from H to A and from D to C, within the coil, forming a current path of DCHA.

In the external circuit, the current flows along B_1RB_2. This current direction remains consistent during the first half-turn of the armature.

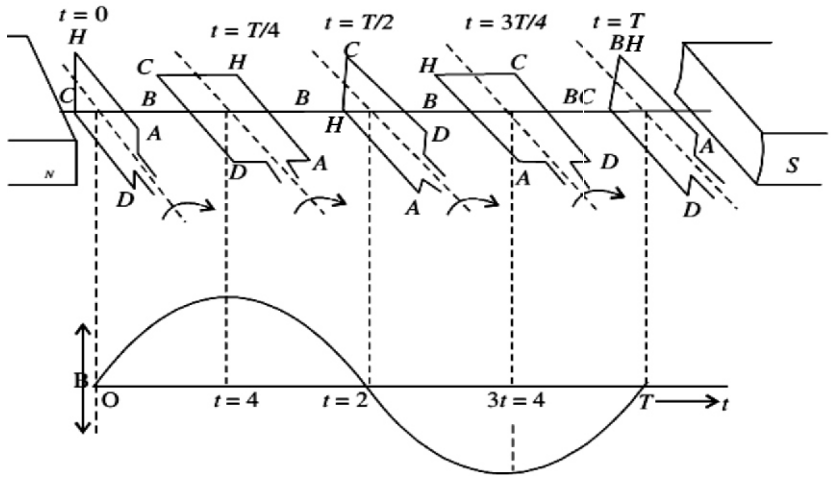


Figure 19.13(a) armature coil AHCD

However, during the second half-revolution, wire AH moves upward, and wires CD move downward, leading to a reversal in the direction of induced current within the coil to AHCD. In the external circuit, the direction changes to B_2RB_1. Therefore, the direction of the induced emf and the current alternates after every half revolution in the external circuit as well. Consequently, the current produced switches direction in each cycle.

When the coil is rotated, a motional electromotive force (emf) is generated in each of its sides, but these emfs have opposite directions because the two sides are moving in opposite directions relative to the magnetic field. However, the other two sides of the coil are moving in the same direction with respect to the field.

Consequently, emfs are induced in these sides in the same direction, leading to a mutual cancellation effect. As a result, the total emf induced in the coil is given by

$$\varepsilon = 2vBNlSin\omega t$$

Where v = linear velocity, B = magnetic field, N = Number of turns, L = length of coil and ω = angular velocity.

Since each particle of the sides AH and CD rotates in a circle of radius equal to the width of the coil b.

$$v = b/2\ \omega$$

$$\varepsilon = 2\ b/2\ \omega\ B\ sin\omega\ t$$

$$\varepsilon = ANB\ \omega\ sin\omega\ t$$

$$\varepsilon = \varepsilon_0\ sin\omega\ t$$

Where ε_0 =ANB ω is the maximum or peak value of the emf which depends on the area and number of the coil, intensity of the field and speed of rotation.

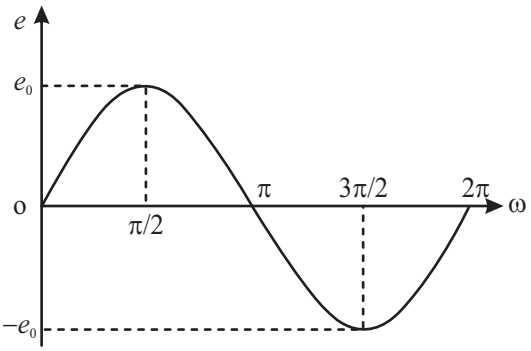


Figure 19.13(b) armature coil AHCD

The expression for the generated emf can also be expressed in terms of the frequency "f" i-e number of turns per second.

$$\varepsilon = \varepsilon_0 \text{Sin} 2\pi ft$$

Figure 19.13(b) shows the variation of emf versus (ω). It shows that direction of emf is reversed after every half revolution of the coil.

Self-Assessment Questions:

1. How does a rotating magnetic field contribute to the rotation of the motor's rotor?
2. Discuss common methods used for starting AC motors.
3. Provide examples of applications where AC motors are commonly used.

Worked Example 19.6

Find the frequency of an AC generator if the voltage output is given as ξ_0 = 200 V,

Solution:

Step 1: Write down the known quantities and quantities to be found.

ξ = 100 V, t = 0.005 seconds.

Step 2: Write down the formula and rearrange if necessary.

$$\varepsilon = \varepsilon_0 \text{Sin} 2\pi ft$$

Step 3: Substitute the known values:

100 = 200 sin (2πf × 0.005)

or

sin (2πf × 0.005) = 0.5

Step 4: To find the frequency, you need to find the value of f that makes the sine function equal to 0.5. This occurs when sin(π/6) = 0.5, which is equivalent to π/6 radians.

So, 2πf × 0.005 = π/6

Solve for f:

f = π / (2 × 0.005 × 6)

f = 52.36 Hz

Result: The frequency of the AC generator is approximately 52.36 Hz.

19.7.1 A.C Motor:

An AC (Alternating Current) motor is a device designed to convert electrical energy into mechanical energy by using alternating current. There are various types of AC motors, but the main features and components are generally consistent across different designs. The primary components as shown in figure and their roles in an AC motor are as follows:

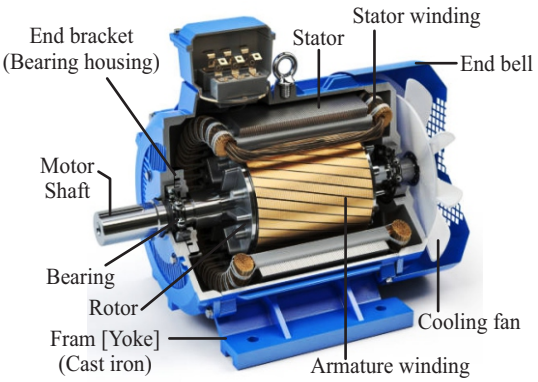


Figure 19.14 AC motor

Stator:

The stator is the stationary part of the motor and contains the primary windings. Its main role is to produce a rotating magnetic field when AC voltage is applied. The rotating magnetic field interacts with the rotor to induce motion.

Rotor:

The rotor is the rotating part of the motor. It can be of different types, such as squirrel-cage or wound rotor, depending on the motor design. When the stator produces a rotating magnetic field, the rotor experiences a torque due to the interaction with the field. This torque causes the rotor to rotate and generate mechanical output.

Bearings:

Bearings are essential components that support and allow the rotor to rotate within the stator. They reduce friction, enabling smooth and efficient operation of the motor.

Shaft:

The shaft is connected to the rotor and extends beyond the motor housing. It transfers mechanical energy to the outside world, allowing the motor to perform useful work when connected to a load.

Cooling System:

Many AC motors incorporate cooling systems, such as fans or fins, to dissipate heat generated during operation. Efficient cooling is important to prevent overheating and prolong the motor's lifespan.

The main features and components work together to enable the AC motor to function. When AC voltage is applied to the stator windings, a rotating magnetic field is created, which exerts a torque on the rotor. The rotor's rotation results in mechanical work being performed by the motor, and the motor can be used to drive various mechanical devices, such as fans, pumps, conveyor belts, and more.

The efficiency and performance of an AC motor depend on the design and quality of its components, the load it drives, and its operating conditions. Proper maintenance and control mechanisms, such as speed control, are often used to optimize the motor's performance in different applications.

19.6.2 The production of back emf in electric motors:

The concept of back electromotive force (back EMF) in electric motors is fundamental to understanding motor operation and efficiency. Back EMF is a self-generated electromotive force that opposes the applied voltage in a motor. It plays a crucial role in motor behavior, especially in limiting current and regulating speed. Here's a detailed explanation of the concept of back EMF in electric motors:

Electromagnetic Induction:

Back EMF is a consequence of Faraday's law of electromagnetic induction, which states that a change in magnetic flux through a coil of wire induces an electromotive force (EMF) in that coil. In an electric motor, the coil of wire is typically part of the rotor or armature.

Rotating Magnetic Field:

When an electric motor is powered, the stator (the stationary part of the motor) generates a rotating magnetic field by applying an alternating current (AC) voltage to its windings. In a direct current (DC) motor, the commutator and brushes create a changing magnetic field.

Rotor or Armature Interaction:

The rotor (or armature) in the motor is mounted within the magnetic field created by the stator. As the rotor rotates, it cuts through the magnetic field lines, causing a change in magnetic flux within the coils of wire on the rotor.

Back EMF Generation:

The change in magnetic flux through the rotor windings induces a voltage, which is referred to as back EMF. The direction of the induced voltage is in opposition to the applied voltage. In other words, it generates a voltage that resists the flow of current through the motor windings.

Current Regulation:

Back EMF has a critical function in motor operation. As the motor speeds up, the back EMF increases. This increase in back EMF results in a decrease in the net voltage across the motor windings. Consequently, the current flowing through the motor decreases. This self-regulation mechanism is essential for preventing the motor from drawing excessive current, overheating, and potentially damaging itself.

Effect on Motor Speed:

The relationship between back EMF and motor speed is inverse. When the motor operates at a higher speed, the back EMF is greater, and the current is reduced, helping maintain a consistent speed. Conversely, at lower speeds or when subjected to a higher mechanical load, the back EMF decreases, allowing more current to flow and providing the necessary torque to overcome the load.

In summary, the concept of back EMF in electric motors is the self-induced voltage that opposes the applied voltage and helps regulate the current and speed of the motor. It is a critical factor in ensuring the safe and efficient operation of electric motors, and understanding it is essential for motor design, control, and performance analysis.

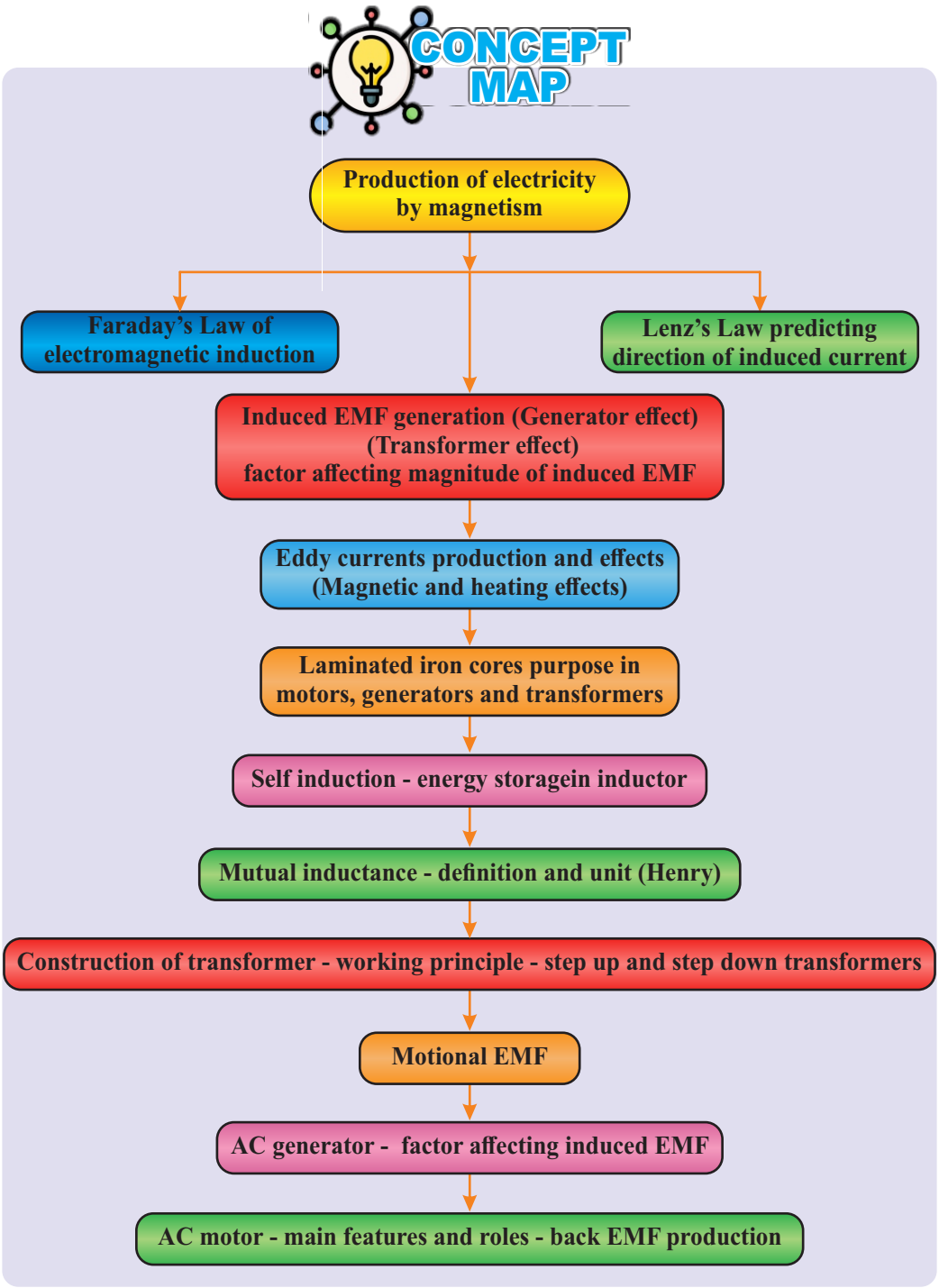


Self-Assessment Questions:

1. Discuss how the production of back EMF aligns with the principle of energy conservation.
2. What is the primary purpose of the back EMF in electric motors?



- ✓ When a current flows through the conductor magnetic field is produced around it.
- ✓ The relative motion of coil or magnet will produce induced emf. If the motion of magnet and coil is stopped induced current ceases to exist.
- ✓ The magnitude of the induced emf depends on the change of magnetic flux, number of turns in the coil, area of the coil, angle between magnetic field and coil.
- ✓ When the magnetic flux $\Phi_B$passing through a closed conducting loop changes over time, it results in the generation of a current and an electromotive force (emf) within the loop.
- ✓ Lenz's law states that ***"The direction of induced emf in a circuit is always such that it opposes the cause which produces it"***.
- ✓ Lenz's law is in accordance with the law of conservation of energy.
- ✓ If a current I in a coil changes with time, an emf is induced in the coil.This self-induced emf is given by $\varepsilon = -L\ \Delta I/\Delta t$
- ✓ An inductor stores energy in the form of a magnetic field when current flows through it, and this stored energy is proportional to the square of the current and the inductance of the inductor.
- ✓ If primary and secondary coils are near each other, a changing current in either coil can induce an emf in the other is known as mutual inductance.
- ✓ A transformer is an electrical device that transfers energy between two or more circuits using electromagnetic induction. It changes voltage levels while keeping power conservation and efficiency intact. The transformer has two or more coils, called windings, connected by a magnetic core, which allows it to increase (step-up) or decrease (step-down) voltage.
- ✓ Motional emf is the electromotive force induced in a conductor moving through a magnetic field, resulting from the relative motion between the conductor and the magnetic field lines.
- ✓ An AC generator converts mechanical energy into electrical energy in the form of alternating current.
- ✓ An AC motor converts electrical energy into mechanical energy.


CONCEPT MAP
Production of electricity by magnetism
Faraday's Law of electromagnetic induction
Lenz's Law predicting direction of induced current
Induced EMF generation (Generator effect) (Transformer effect) factor affecting magnitude of induced EMF
Eddy currents production and effects (Magnetic and heating effects)
Laminated iron cores purpose in motors, generators and transformers
Self induction - energy storagein inductor
Mutual inductance - definition and unit (Henry)
Construction of transformer - working principle - step up and step down transformers
Motional EMF
AC generator - factor affecting induced EMF
AC motor - main features and roles - back EMF production



Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. Electricity production by magnetism involves
   (a) the conversion of chemical energy to electrical energy.
   (b) the use of static electricity for power generation.
   (c) the movement of conductors within magnetic fields.
   (d) the absorption of solar energy by photovoltaic cells.
2. Induced e.m.f can be generated by relative movement, known as the generator effect, and by changing magnetic fields, known as the transformer effect.
   (a) True in transformers only.
   (b) False in all cases.
   (c) Applicable to both cases.
   (d) True only for static magnetic fields.
3. The magnitude of induced e.m.f. increases with
   (a) the decrease in the speed of the conductor's motion.
   (b) the decrease in the strength of the magnetic field.
   (c) the increase in the number of turns in the coil.
   (d) the reduction in the area of the coil.
4. Faraday's law of electromagnetic induction states
   (a) that current in a circuit always opposes the change in magnetic flux.
   (b) that the induced e.m.f. is directly proportional to the rate of change of magnetic flux.
   (c) that a changing electric field produces a magnetic field.
   (d) that magnetic fields can only be produced by electric currents.
5. Lenz's law predicts the direction of an induced current
   (a) to conserve electric charge.
   (b) by stating that the induced current will oppose the change causing it.
   (c) by following the direction of the applied magnetic field.
   (d) based on the magnitude of the electric field.
6. Eddy currents produce:
   (a) static electricity in conductive materials.
   (b) alternating magnetic fields with no heating effects.
   (c) magnetic fields that oppose the inducing field and cause heating in conductors.
   (d) uniform electric fields that have no impact on power generation.
7. Laminated iron cores are used in electric motors and transformers to
   (a) reduce the weight of the device.
   (b) enhance the mechanical strength.
   (c) minimize eddy current losses and reduce heating.
   (d) increase the magnetic permeability of the core.

8. Self-induction is the phenomenon where
   (a) a changing electric field induces a magnetic field.
   (b) a changing magnetic field within a coil induces an emf in the same coil.
   (c) a constant magnetic field induces a constant emf.
   (d) two coils induce emf in each other through mutual induction.
9. An inductor stores electric potential energy in:
   (a) the electric field around it.
   (b) the magnetic field within its coil.
   (c) the capacitance of its windings.
   (d) the heat generated by its resistance.
10. Transformers work on the principle of
   (a) converting direct current (DC) to alternating current (AC).
   (b) electromagnetic induction between primary and secondary coils.
   (c) generating electricity through chemical reactions.
   (d) using permanent magnets to maintain a constant voltage.

Section (B): CRQs (Short Answered Questions):

1. Explain Faraday's law of electromagnetic induction.
2. State Lenz's law and discuss its significance in the context of electromagnetic induction.
3. How does a step-up transformer differ from a step-down transformer?
4. Describe the principle of operation of an AC generator.
5. Define self-induction and explain how it occurs in a coil.
6. Define mutual induction and provide an example of a system exhibiting mutual induction.
7. How does the arrangement of coils influence the degree of mutual induction between them?
8. Define motional electromotive force (emf)
9. Differentiate between motional emf and electromagnetic induction in terms of their fundamental principles.
10. Differentiate between AC (alternating current) and DC (direct current) in the context of long-distance power transmission.

Section (C): ERQs (Long Answered Questions):

1. How does the strength of the magnetic field affect the induced electromotive force (emf) in a coil according to Faraday's law? b. Explain the relationship between the magnetic field strength and the magnitude of the induced emf.
2. In Faraday's law, why is the area of the coil an important factor in determining the induced emf?
3. How does changing the orientation of a coil with respect to the magnetic field influence the induced emf? Also describe the role of the number of turns in a coil in the context of induced emf. How does increasing the number of turns in a coil affect the induced emf, and why?
4. According to Faraday's law, why is the rate of change of magnetic flux crucial for inducing emf? Discuss how the speed of motion or change in magnetic field influences the induced emf.

5. How does the resistance of the conductor impact the flow of induced current in a circuit? In what situations might the resistance of the conductor become a significant factor in induced emf?
6. Explain the relationship between the frequency of the changing magnetic field and the induced emf. How does the frequency of the AC source affect the generation of induced emf in a coil?
7. Discuss how the material properties of a coil, such as its conductivity, can influence induced emf. In what ways might the type of material used in a coil affect its response to changing magnetic fields?
8. Why is the presence of a core material significant in transformers concerning induced emf? Explain how the type and properties of the core material can impact the efficiency of a transformer.
9. Why are slip rings used in AC generators, and how do they differ from commutator used in DC generators.
10. Why a commutator is necessary in DC motors, and how does it facilitate the continuous rotation of the motor?
11. What are the common applications of DC motors in everyday devices?

Section (D): Numerical:

1. Two coils are placed adjacent to each other, and a change in current in the first coil induces an emf of 0.5 V in the second coil. If the mutual inductance is 0.2 H, calculate the rate of change of current in the first coil. **(2.5A/s)**
2. A coil with an inductance of 0.5 H experiences a rate of change of current of 2 A/s. Calculate the induced EMF in the coil. **(-1V)**
3. Two coils are placed close to each other. If a change in current of 3 A/s in the first coil induces an EMF of 4 V in the second coil, calculate the mutual inductance. **(1.33H)**
4. A transformer has 200 turns in the primary coil and 400 turns in the secondary coil. If the primary voltage is 120 V, calculate the secondary voltage for a step-up transformer. **(240V)**
5. An inductor with an inductance of 0.02H has a current flowing through it of 2 A. Calculate the energy stored in the inductor. **(0.04J)**
6. A coil stores energy in form of electric potential energy of 0.2J when it carries a current of 2A. Calculate the inductance of coil. **(0.1H)**
7. An AC generator produces an alternating current with a maximum voltage of 240 V. If the frequency of the generated AC is 50 Hz, calculate the peak value of the voltage. **(339V)**
8. A transformer has 1000 turns in its primary coil and 200 turns in its secondary coil. If the primary voltage is 120 V, calculate the secondary voltage for a step-down transformer. **(24V)**
9. A conductor of length 0.4 m moves at a velocity of 5 m/s perpendicular to a magnetic field of 0.3 T. Calculate the motional EMF induced in the conductor. **(0.6V)**

Unit 20 AC Circuit

Teaching Periods 15 **Weightage %** 10



The Jamshoro grid station is the sole transit point for affordable power from K2, K3, China Power Hub, and Hubco to meet northern demand. It is also the only transit point for evacuating wind power from Jhimpir/Gharo. Grid stations, are critical components for the distribution and utilization of AC (Alternating Current) circuits.

In this unit student should be able to:

- Define the terms time period, frequency, instantaneous peak value and root mean square value of an alternating current and voltage.
- Represent a sinusoidal alternating current or voltage by an equation of the form $x = x_o \sin \omega t$.
- Describe the phase of A.C and how phase lags and leads in A.C Circuits.
- Explain the flow of A.C through Resistors, Calculate the resistance of resistors
- Construct phasor diagrams and carry out calculations on circuits including resistive components
- Explain the flow of A.C through capacitors
- Calculate the reactance of capacitors
- Construct phasor diagrams and carry out calculations on circuits including reactive components
- Explain Ac through inductors.
- Identify inductors as important components of A.C circuits termed as chokes
- Construct phasor diagrams and carry out calculations on circuits including inductive components
- Describe impedance as vector summation of resistances and reactance
- Construct phasor diagrams and carry out calculations on circuits including resistive reactive and inductive components in series and parallel.
- Solve the problems using the formulae of A.C Power.
- Explain resonance in an A.C circuit and carry out calculations using the resonant frequency formulae.
- Describe that maximum power is transferred when the impedances of source and load match to each other.
- Illustrate the principle of metal detectors used for security checks.
- State the principle of electro-cardiograph in medical diagnostic.
- Describe the importance of oscillator circuit as broadcaster of radio waves.
- Describe the principle of resonance in tuning circuits of a radio.

Introduction:

Alternating current is the back bone of electrical engineering and physics. In addition, we will explore the dynamic nature of AC current, which is quite different when comparing with its counterpart, Direct Current (DC).

There is great contribution of many scientists, like Nikola Tesla and George Westinghouse in better understanding of alternating current. Their groundbreaking work in the late 19th and early 20th centuries laid the foundation for the widespread use of AC power, revolutionizing the way electricity is generated, transmitted, and utilized.

In this chapter we will discuss essential concepts such as frequency, amplitude, and the sinusoidal waveform that characterizes AC current. We will use the mathematics behind AC circuits and study the principles governing the behavior of resistors, capacitors, and inductors when connected with alternating current source. Also, the practical applications of AC current, from the power distribution grid that lights up our cities to the metal detector that have become integral to our daily lives.

20.1 Alternating Current

A graphical representation, such as the one depicted in Figure 20.1, shows the nature of an alternating current. This graph visually demonstrates the cyclic variation of the current, depicting a pattern where, during one half of the cycle, the current is positive, while during the other half, it is negative. Consequently, the current alternates its direction in the wires it traverses, flowing first in one direction and then in the opposite direction.

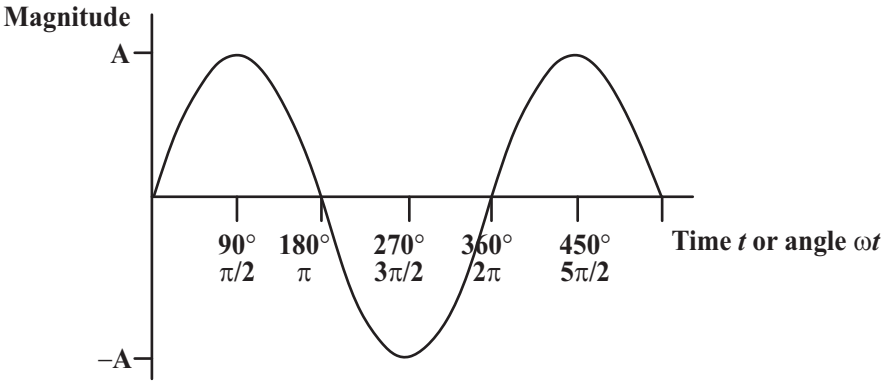


Figure 20.1 A graphical representation of sinusoidal alternating current.

In the context of utilizing household appliances connected to the mains, the current continuously oscillates back and forth along the wires connecting the user to the power station generating the electricity. At any given moment, the current possesses a specific magnitude and direction, as indicated by the graph.

Alternating current (AC) varies in a sinusoidal pattern and follows the shape of a sine wave. This pattern is characterized by a smooth, periodic oscillation that changes direction between positive and negative. While any current that alternates direction could be considered AC, the term typically refers to currents that exhibit this regular, sinusoidal pattern. This sinusoidal nature is fundamental to how AC is generated and utilized in power systems.

20.1.1 AC terminologies

Cycle:

One complete set of both positive and negative values of an alternating quantity is known as a cycle as shown in figure 20.2.

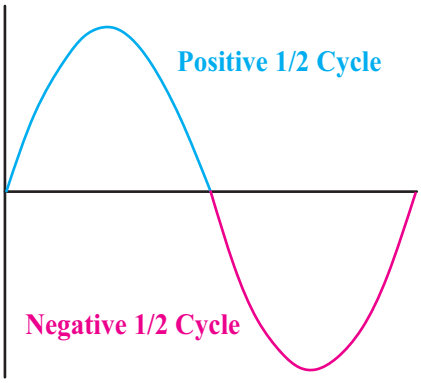


Figure 20.2 One cycle of ac current

Time Period (T):

The time period of an alternating current (AC) or voltage wave is the duration it takes to complete one full cycle. It is typically denoted by the symbol T. The relationship between time period and frequency (f) is given by the equation: $= \frac{1}{f}$, where T is in second.

Frequency (f):

It refers to the number of complete cycles of the current waveform that occur in one second. It is a measure of how many times the direction of the current reverses per second. Frequency is typically expressed in hertz (Hz), where one hertz equals one cycle per second. In practical terms, the frequency of AC current determines how rapidly the current alternates its direction and is crucial for the operation of electrical devices that rely on AC power. The relationship between frequency and time period is expressed as $f = \frac{1}{T}$

Instantaneous Peak Value:

The instantaneous peak value refers to the maximum amplitude of the alternating current or voltage at any specific point in time. For a sine wave, this is the highest positive or negative value reached during a cycle. It is usually denoted as I_{peak} for current or V_{peak} for voltage.

Root Mean Square (rms) Value:

The average value of voltage or current is not used in electric power calculations. The reason being that the AC cycle consists of positive and negative half cycles so the average value is zero. However, there exists a mathematical relation between the peak value V_0 of alternating voltage and a direct current (d.c) voltage that yields an equivalent average electrical power. This constant d.c. voltage is referred to as the root-mean-square (r.m.s) value of the alternating voltageV_{rms}. The d.c voltage is approximately 70% of V_{peak} as shown in figure 20.3. For a sinusoidal wave, the rms value is approximately 0.707 times the peak value. Mathematically,

$V_{rms} = 0.707 \times V_{peak}$ for voltage

and $I_{rms} = 0.707 \times I_{peak}$ for current.

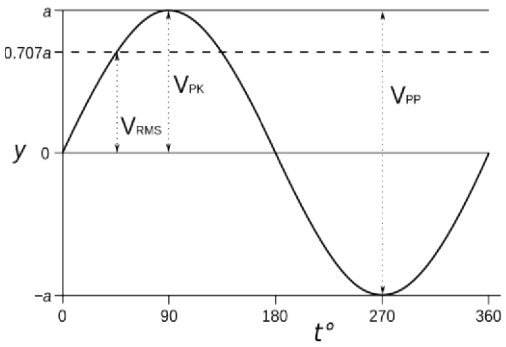


Figure 20.3 Cycle of ac current

20.1.2 Alternating Current OR voltage equation:

When a coil undergoes constant rotation at a uniform angular velocity within a magnetic field, it generates an alternating voltage. Alternating Current (AC) or Voltage refers to the type of electrical flow or potential difference that periodically changes direction and magnitude over time. The fundamental equation describing AC voltage or current in a sinusoidal waveform is:

$V = V_0 Sin\omega t$..... (20.1)

$I = I_0 Sin\omega t$..... (20.2)

Where,

V, I represent the instantaneous value of the voltage and alternating current respectively

V_0, I_0 is the amplitude of the waveform, which represents the maximum value of alternating voltage and alternating current respectively

ω is the angular frequency, given by $\omega = 2\pi f$

Worked Example 20.1

If an alternating current is flowing through a circuit with an amplitude of 5 amps, with frequency of 60 Hz, write its equation:

Solution:

Step 1: Write the known quantities and point out quantities to be found

$Vo = 5A$,

$f = 60\ Hz$

Step 2: Write the formula and rearrange if necessary

$$I = I_0 Sin\omega t$$

Step 3: Now, substitute the values into the formula:

$$I = 5\ Sin2\pi \times 60 \times t$$

Result: The sine function will oscillate between -5 and 5 Amps.

20.1.3 Phase of AC:

The phase difference between two waves carries more important than their magnitudes. The phase describes the relative positions of these waveforms. Hence in alternating current (A.C.) circuits, the concept of phase refers to the relationship between different waveforms as time passes. Now we discuss different states of phase within AC circuit.

In-phase:

When two waveforms have the same frequency and reach their peak values or zero values at the same time, they are said to be in phase as shown in figure 20.4.

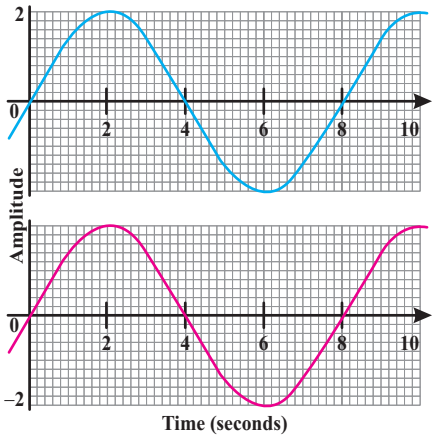


Figure 20.4 Two waves are in a phase

Phase Lag:

If one waveform reaches its peak or zero value after the other, there is a phase difference, and the second waveform is said to lag the first. It is shown by figure 20.5 where a wave shown by dotted line lags the wave shown by solid line. It can be said that the dotted wave form lags the solid line waveform by 90^0 or $\frac{\pi}{2}$ radians.

Phase Lead:

Conversely, if one waveform reaches its peak or zero value before the other, the second waveform is said to lead the first. It is shown by figure 20.5 where a wave shown by solid line leads the wave shown by dotted line by phase difference of 90^0 or $\frac{\pi}{2}$ radians.

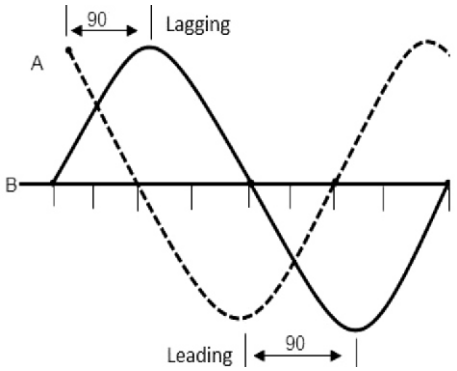


Figure 20.5
The two waves are 90° out of phase.

The representation of phase lead and lag between two alternating quantities is effectively demonstrated by showing the two AC quantities as vectors also known as "Phasor".

Vector Representation of an Alternating quantity:

To solve AC problems effectively, it is beneficial to represent a sinusoidal quantity, such as voltage or current, using a line of specific length that rotates in a counterclockwise direction with the same angular velocity as the sinusoidal quantity. This rotating line is commonly referred to as a phasor.

Let's consider a line denoted as OA, referred to as a phasor, which accurately represents, to scale, the maximum value of an alternating quantity, such as electromotive force (emf). In this representation, OA equals the maximum emf value and rotates counterclockwise at an angular velocity of ω radians per second around the point O, as illustrated in Figure 20.6. An arrowhead is placed at the outer end of the phasor, serving both to indicate the assumed direction of movement and to specify the precise length of the phasor, especially when multiple phasors coincide.

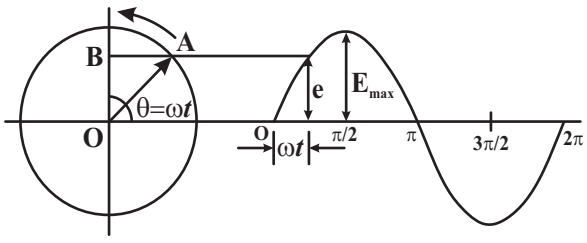


Figure 20.6 The propagation of sinusoidal wave and its vector (phasor) representation

In Figure 20.6, OA represents the phasor after it has rotated through an angle θ, equivalent to ωt, from its initial position when the emf was at its zero value. The projection of OA on the Y-axis, denoted as OB, equals OA multiplied by sin θ, which is equivalent to Emax sin ωt, representing the instantaneous value of the emf, denoted as 'e,' at that particular moment. Therefore, the projection of OA on the vertical axis accurately portrays, to scale, the instantaneous value of the emf.

Self-Assessment Questions:

1. Explain the concept of reactance in AC circuits and how it differs from resistance.
2. How does the phase angle affect the relationship between voltage and current in AC circuits?

20.2 AC through a resistor

Consider a circuit comprising a pure resistor (R) connected to an alternating voltage source, as depicted in figure 20.7(a).

The alternating voltage induces oscillatory motion of free electrons within the resistor, constituting the alternating current. At any given time 't,' the potential difference across the resistor's terminals is expressed as

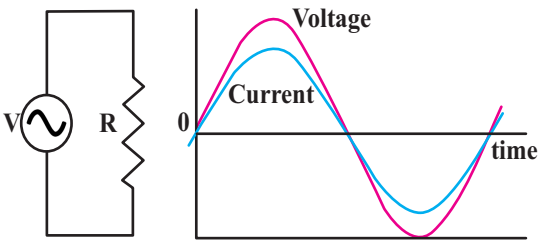


Figure 20.7(a) resistor connected to an AC source and 20.7(b)the voltage and current wave are in a phase

$$V = V_0 sin\omega t$$

where V_0 signifies the peak value of the alternating voltage. The circuit's current (I) is governed by Ohm's law: $I = \frac{V}{R} = \frac{V_0}{R} Sin\omega t$

Or $I = I_0 Sin\omega t$..... (20.3)

The current achieves its maximum value when $Sin\omega t = 1$. From equations (20.2) and (20.3), it is evident that the applied voltage and the current in the circuit are in phase with each other as shown in figure 20.7 (b). This phase relationship is further illustrated by the phasor diagram as shown in figure 20.8.

20.2.3 Phasor diagram of Resistive circuit

The current I_R flowing through a resistor R is in phase with the voltage V_R across the resistor. This alignment can be visually represented on a phasor diagram by depicting a vector (I_R) that coincides with the voltage (V_R)

The power loss in a resistor in an AC circuit is a result of the conversion of electrical energy into heat due to the resistance of the resistor. The instantaneous power in the resistance is can be expressed using the formula

$P = I^2R$..... (20.4)

$P = V_{rms} I_{rms}$..... (20.5)

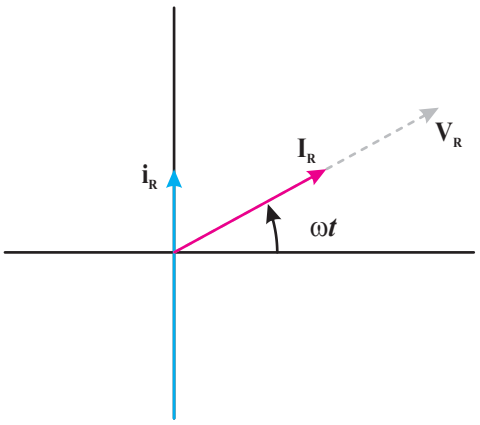


Figure 20.8 phasor diagram of resistive circuit

Worked Example 20.2

A 200Ω resistor is connected with AC supply of 36V with frequency 60 Hz. Calculate
(i) Current flowing through the circuit
(ii) Power dissipated in the resistor.
(iii) Also write equation for current and voltage as a function of time.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$R = 20\Omega$
$V = 36\ Volts$
$f = 60\ Hz$

Step 2: Write the formula and rearrange if necessary

$$I = \frac{V}{R} = \frac{36}{200} = 0.180A$$

Step 3: Power dissipation can be calculate using equation

P = VI = 36 × 0.180 = 6.48 w
We know that ω = 2πf = 2 × 3.14 × 60 = 376.8 Hz
From equations (2) and (3)
$V = V_0 \sin\omega t$
$I = I_0 \sin\omega t$

Step 4: Substituting values of V, I and ω in above equations, we get

V = 36 sin376t
I = 0.180 Sin376t

20.3.1 A.C through capacitors:

When capacitor is connected to a direct current (DC) supply voltage, then capacitor plates gradually accumulate charge until the voltage across the capacitor matches with the source voltage. The capacitor retains this charge as a temporary storage device as long as the applied voltage is sustained. The ability of a capacitor to store charge on its plates is termed capacitance (C)

During the charging process, an electric current flow into the capacitor, causing its plates to acquire an electrostatic charge. This charging current is highest when the capacitor plates are uncharged, decreasing exponentially over time until the capacitor reaches full charge. Once the capacitor is fully charged, it blocks further electron flow onto its plates as they become saturated. Hence it is often said that capacitor blocks direct current after initial transient time.

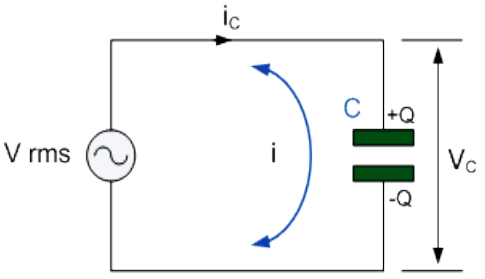


Figure 20.9 capacitor connected to the AC source

Now consider a capacitor connected to the AC source as shown in figure 20.9. during the positive half cycle of alternating voltage, the electrons flow from upper capacitor plate to

the source leaving it as positively charged +Q and source supplies equal number of electrons to the lower plate making it as negatively charged –Q as shown in figure 20.9. During the negative half cycle of alternating voltage, the direction of motion of electrons is also reverted resulting in the capacitor plates becoming charged in the opposite manner and current flows in opposite direction, hence capacitor charges and discharges. In this way the alternating current flows through the capacitor.

If alternating source voltage $V = V_0 sin\omega t$ is applied to the capacitor the charge on any plate of capacitor is given by

$q = CV = CV_0 Sin\omega t$..... (20.6)

From equation (20.6) it is clear that q and V are in phase. Due to the applied current flows through the capacitor as given by

$$I = \frac{\Delta q}{\Delta t} \text{ ..... (20.7)}$$

Substitute equation (20.6) in (20.7), we get

$$I = \frac{CV_0 \Delta Sin\omega t}{\Delta t} \quad \text{OR} \quad I = CV_0 \omega Cos\omega t \text{..... (20.8)}$$

Multiply and divide right side of equation (20.8) by ωc

$$I = \frac{V_0 Sin(\omega t + \frac{\pi}{2})}{1/\omega c}$$

$$I = I_0 \, Sin(\omega t + \frac{\pi}{2}) \text{ ..... (20.9)}$$

Equation (20.9) shows that in pure capacitive circuit the current exhibits sinusoidal variation and it leads the voltage by 90 degrees or $\frac{\pi}{2}$ radians or the voltage lags the current by 90 degrees or $\frac{\pi}{2}$ radians as shown in figure 20.10.

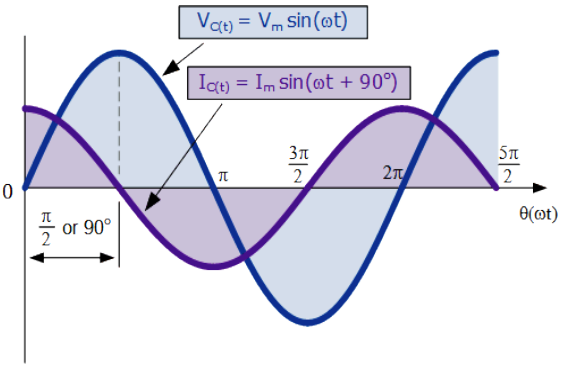


Figure 20.10 Pure capacitive circuit the current varies sinusoidally and leads the voltage by 90 degrees.

The reason is that when a voltage is introduced to an initially uncharged capacitor, the capacitor exhibits low impedance, resulting in the maximum current draw. As the capacitor becomes charged, the current diminishes, causing the voltage across the capacitor to increase. Once the capacitor is fully charged, there is no further current flow, and the voltage attains its maximum level, hence voltage lags the current by 90^0.

20.3.2 Reactance of Capacitor:

In a purely capacitive circuit, capacitive reactance represents the opposition to alternating current flow. Similar to resistance, reactance is measured in Ohms, but it is denoted by the symbol Xc to differentiate it from purely resistive values.

From equations (9), we get

$$I = \frac{V_0}{1/\omega c} \text{..... (20.10)}$$

Similar to ohms law the term in denominator of equation (20.10) represents the opposition to flow of current; in this case it is called reactance of capacitor.

Hence

$$X_c = \frac{1}{\omega c} \ldots\ldots\ldots\ldots (20.11)$$

Since $\omega = 2\pi f$ so

$$X_c = \frac{1}{2\pi f c}$$

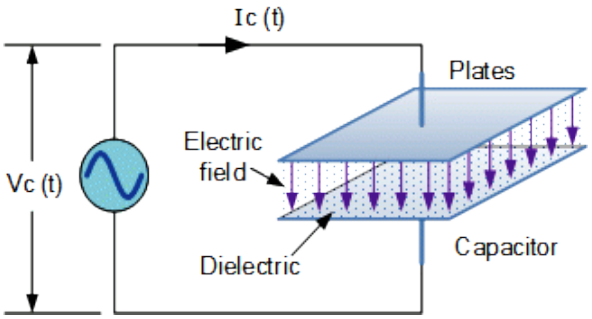


Figure 20.11 Reactance of Capacitor

The capacitive reactance shows an inverse relationship with the frequency of the applied alternating voltage. Consequently, for lower frequencies, the reactance of capacitor increases, while at high frequencies, the reactance decreases.

20.3.3 Phasor diagram of capacitive circuit:

As mentioned in section 20.3 that in an AC circuit, a capacitor experiences a phase shift between the voltage across it and the current flowing through it. The phasor diagram is a graphical tool that illustrates the relationship between voltage and current in a capacitor using vectors, known as phasor. In the phasor diagram shown in Figure 20.12, V represents the voltage across the capacitor, while I denote the current flowing through it. The angle $\emptyset = 90^o$ indicates the phase angle by which the current leads the voltage in a capacitor.

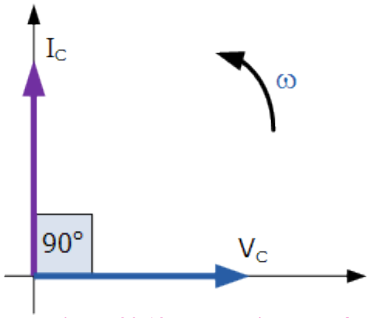


Figure 20.12 Phasor diagram of capacitive circuit

Self-Assessment Questions:

1. Why is the root mean square (RMS) value used to represent AC voltage and current instead of instantaneous peak values?
2. How does the phase angle affect the relationship between voltage and current in AC circuits?

Worked Example 20.3

A 200Ω resistor is connected with AC supply of 36V with frequency 60 Hz. Calculate
(i) Current flowing through the circuit (ii) Power dissipated in the resistor.
(iii) Also write equation for current and voltage as a function of time.

Solution:

Step 1: Write the known quantities and point out quantities to be found

R = 200Ω
V = 36 Volts
f = 60 Hz

Step 2: Write the formula and rearrange if necessary

$$X_c = \frac{1}{2\pi fc}, \quad I_{rms} = \frac{V_{rms}}{X_c}$$

Step 3: Now, calculate the current: =

Xc =1/2πfc = 1/ (2×3.14×50×10×10 ^ (-6)) = 3183.1 Ω

I_{rms}= V_{rms}/X_c = 220/3183.1=0.069A

Step 3: First calculate the maximum current and voltage by using

I_0= √2I=0.097A

V_0=√2 V=311.1

V=311.1 Sin (314 t)

I=0.097 Sin (314 t+π/2)

20.4 A.C through inductor:

An inductor is a passive electrical component which can be formed by wounding a conducting wire over an insulating object, such as pencil. The primary purpose of an inductor is to oppose changes in current. It resists the flow of alternating current (AC) and stores energy in its magnetic field during the on-time of the AC cycle and releases it during the off-time. Solenoid is an example of inductor.

Consider an inductor in form of solenoid connected with A.C source as shown in figure 20.13.

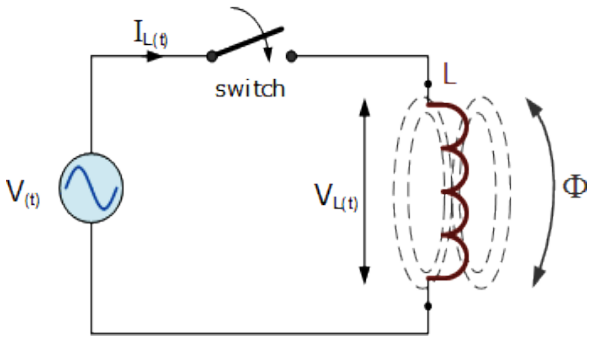


Figure 20. 13 A.C through inductor.

When the switch is closed current start to flow through the inductor, notice that the magnitude and direction of current is changing hence associated magnetic field also varies due to which an induced emf is set up in the inductor so as to oppose the change in accordance with the Lenz law. The magnitude of induced emf is given by

$$\varepsilon = L\frac{\Delta I}{\Delta t}$$

Therefore, to sustain the current, the applied voltage must match the back electromotive force (EMF). Thus, the magnitude of voltage supplied to the coil is expressed as

V= $L\frac{\Delta I}{\Delta t}$…….. (20.12)

The alternating voltage produces a sinusoidal current given by

$$I = I_0 Sin\omega t$$

Hence equation (20.12) becomes

$V = L\frac{I_0 \Delta sin\omega t}{\Delta t}$ Or

$V = \omega L I_0 \, Cos\omega t$

Where $\omega L I_0 = V_0$

$V = V_0 \, Cos\omega t$..... (20.13)

Using that $Cos\omega t = Sin\,(\omega t + \frac{\pi}{2})$, equation (13) becomes

$V = V_0 \, Sin\,(\omega t + \frac{\pi}{2})$..... (20.14)

From equation (20.14) it is clear that in case of inductor connected in A.C circuit the voltage leads the current by $\frac{\pi}{2}$ radians as shown in figure 20.14.

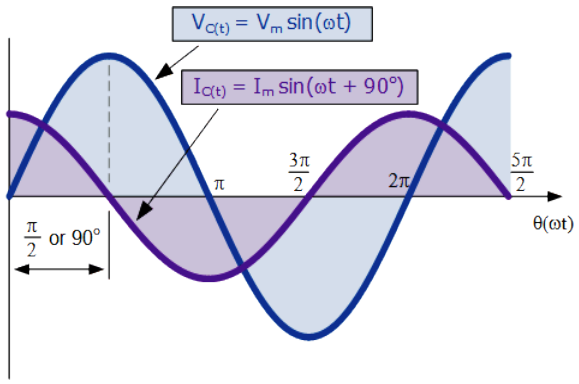


Figure 20.14 Voltage leads current by 90

The reason for voltage leading in an inductive circuit is due to the generation of induced electromotive force (EMF), commonly known as back EMF, when an alternating voltage is applied to the inductor. This back EMF appears instantaneously and induces a counter-current flow, introducing a time delay, typically in the order of milliseconds. As a result, there is a brief time lag for the circuit current to overcome this opposing current and attain its maximum value. That is why the voltage manifests first followed by the appearance of the current after a short interval. Hence the voltage leads the current or the current lags behind the voltage in an inductive AC circuit, draw I_L behind V_L by a phase angle of 90 degrees as shown by phasor diagram in figure 20.15.

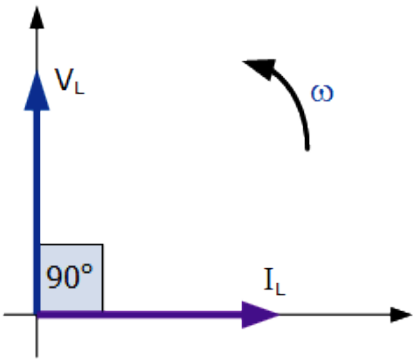


Figure 20.15
Phasor diagram of inductive circuit

Choke coil:

A choke coil, commonly known as choke, is an inductor used in electronic circuits. It is used to block high frequencies above a certain frequency, while allowing the direct current to pass through. It is formed within a core shaped like a donut, around which an insulated coil is wound as shown in figure 20.16. This important electrical component is only used in A.C circuits. Its ability to impede changes in current makes it valuable in applications where a smooth and noise-free DC signal is required. Choke coils find applications in various electronic devices, including power supplies for audio amplifiers, radio frequency (RF) circuits, and other electronic systems where filtering and noise suppression are essential.



Figure 20.16 Chock coil

20.4.1 Inductive Reactance:

The term inductive reactance refers to the opposition that an inductor offers to the flow of alternating current. It is denoted by X_L. It is measured in ohms. The inductive reactance can be calculated by following the method as described in section 20.3.1.

The formula for inductive reactance is given by:

$X_L = \omega L$ Or $X_L = 2\pi f L$..... (20.15)

where: X_L is the inductive reactance,

f is the frequency of the AC signal,

L is the inductance of the inductor.

Equation (20.15) shows that inductive reactance is directly proportional to the frequency of the AC signal. As the frequency increases, the inductive reactance also increases.

Worked Example 20.4

Suppose we have an inductor in form of a coil with an inductance of 0.02henries (H) and connected in series with an AC signal having frequency of 50 Hz. Calculate the inductive reactance of coil.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$L = 0.02H$

$f = 50Hz$

$X_L = ?$

Step 2: Write the formula and rearrange if necessary

$$X_L = 2\pi f L$$

Step 3: Substituting the give values in formula, we get

$X_L = 2 \times 3.14 \times 50 \times 0.02$

$X_L = 6.28\ \Omega$

20.5 RLC Circuits:

An RLC circuit is an electrical circuit consisting of a resistor (R), inductor (L), and capacitor (C), connected in series or parallel. RLC circuits are fundamental in electronics for filtering signals, tuning circuits, and resonant applications due to their ability to manipulate frequencies and impedance.

Impedance triangle:

A circuit containing resistor, inductor and capacitor offer opposition to flow of current due to all these circuit elements known as impedance. It is denoted by Z. The impedance triangle is shown in figure 20.17.

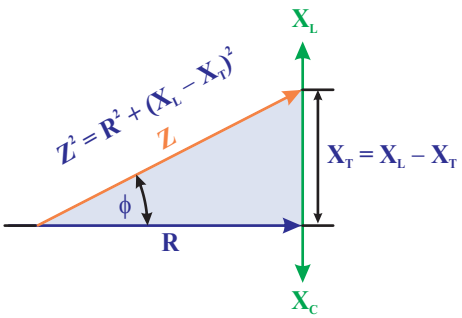


Figure 20.17 impedance triangle

To draw an impedance triangle, represent the resistance (R) as the horizontal side of the triangle.

Represent the inductive reactance (X_L) or capacitive reactance (X_C) as the vertical side of the triangle. The direction (up or down) depends on whether it's inductive or capacitive reactance.

Now the hypotenuse of the triangle represents the impedance (Z). From the right-angle triangle impedance can be calculated by using following relation

$Z = \sqrt{R^2 + X_T^2}$..... (20.16)

Where Z is impedance, R is resistance in a circuit and $X_T^2$is sum of capacitive and inductive reactance of circuit

20.5.1 AC through RC Series circuits:

Consider a circuit containing resistor and capacitor connected in series with an alternating voltage source as shown in figure 20. 16. Due to series circuit same current will flow through each circuit element. While voltage across resistor will be V_R and V_c across capacitor.

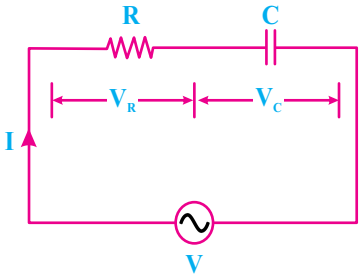


Figure 20.18 RC series circuit

The phasor diagram is drawn taking current as reference direction as shown in figure 20.18. As discussed in section 20.3.3 the voltage lags the current by $\frac{\pi}{2}$ radians in case of capacitor so it is drawn perpendicular to the current phasor in figure 20.19. In RC series circuit the voltage and current are in phase when considering resistor component of circuit.

From the phasor diagram, using pythagoras theorem

$$V = \sqrt{V_R^2 + V_c^2} \text{..... (20.17)}$$

Substituting volatage across resistor = IR and across capacitor =IXc in eqaution (16), we get

$$V = I\sqrt{R^2 + X_c^2} \text{..... (20.18)}$$

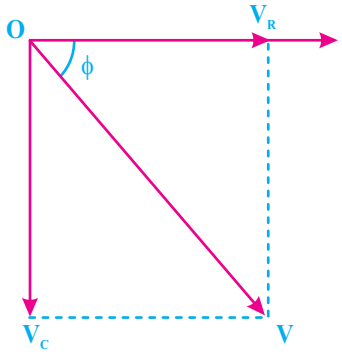


Figure 20.19
Phasor diagram of RC series circuit

Comparing eqautions (20.16) and (20.18), we find that qauntity $\sqrt{R^2 + X_c^2}$ represents the impedence of RC series circuit. $Z = \sqrt{R^2 + X_c^2}$

Phase angle between current and voltage as shown in figure 20.19 is calculated by

$$tan\emptyset = \frac{V_c}{V_R} = \frac{X_c}{R} \qquad \text{Or} \quad \emptyset = tan^{-1}\left(\frac{X_c}{R}\right)$$

20.5.2 AC through RL Series circuits:

Next in order to calculate the impedence of RL series circuit, consider a circuit containing indcutor and resistor connected in series with an alteranting voltage as shown in figure 20.20(a). The volatge across resistor and indcuitor in V_R and V_L respectively.

As discussed in section 20.3.6 when an indcutor is connected with alternating voltage then due to generation of back EMF current lag the voltage by 90^0 it is repredented by line perpendciular to the referenece phasor i-e current I as shown in fgiure 20.20(b). Also as discussed in section 20.31 in resisitor the current and voltage are in phase. In phasor diagram it is V_R line.

To calculate the impedance of RL circuit, apply pythagoras theorem on phasor diagram from figure 20.18 (b)

$$V = I\sqrt{R^2 + L_c^2}$$

Where $Z = \sqrt{R^2 + L_c^2}$ represents the impedance of the RL circuit.

Phase angle between current and voltage as shown in figure 20.20(b) is calculated by

$tan\emptyset = \frac{X_L}{R}$ Or $\emptyset = tan^{-1}\left(\frac{X_L}{R}\right)$

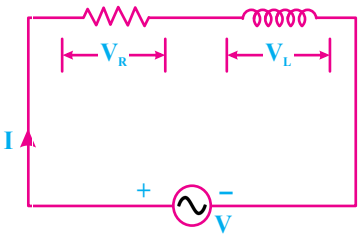


Figure 20.20 (a) Inductor and resistor connected in series

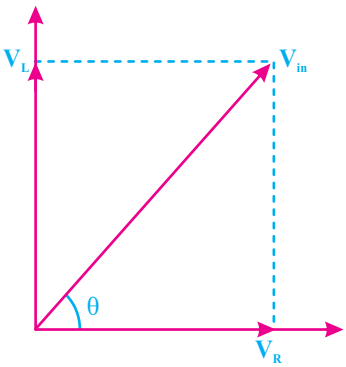


Figure 20.20 (a) Inductor and resistor connected in series

Worked Example 20.5

In RC circuit a capacitor of 50μF and resistor of 100 Ω are connected in series with alternating voltage source of 220V and frequency of 50Hz. Find the (a) current in the circuit (b) phase angle between voltage and current.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$C = 50\times10^{-6}F$ $R = 100\ \Omega$, $V = 220Volts$, $f = 50Hz$

Step 2: Write the formula and rearrange if necessary

$$X_c = \frac{1}{2\pi fc} =$$

Step 3: Substituting the give values in formula, we get

The capacitive reactance $= X_c = \frac{1}{2\pi fc} = \frac{1}{6.28\times50\times50\times10^{-6}} = 636.9H$

Impedance $= Z = \sqrt{R^2 + X_c^2} = \sqrt{100^2 + 636.9^2} = 644.7\Omega$

Current $= I_{rms} = \frac{V_{rms}}{Z} = \frac{220}{636.9} = 345mA$

Phase angle $= \emptyset = tan^{-1}\left(\frac{X_c}{R}\right) = tan^{-1}\left(\frac{636.9}{100}\right) = 81.04^0$

20.5.3 RLC Series AC Circuit:

An RLC series circuit consists of a resistor (R), inductor (L), and capacitor (C) connected in series to an AC voltage source as shown in figure 20.21. In such a circuit, the electrical components influence the behavior of the current flowing through the circuit.

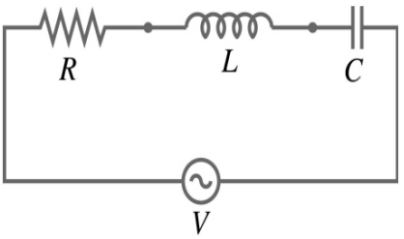


Figure 20.21 RLC Series AC circuit

Since all three circuit elements are connected in series so same current will flow through them. Hence to analyze the different properties of circuit we draw a phasor diagram of RLC series circuit as shown in figure 20.22.

In section 20.4 and 20.5 the capacitive and reactance were determined as $X_c = \frac{1}{2\pi fc}$ and $X_L = 2\pi fL$. for smaller frequency capacitive reactance is much greater than inductive reactance i-e $X_C > X_L$ causing the circuit to exhibit behavior akin to an RC circuit. For larger frequency inductive reactance dominates over capacitive reactance i-e $X_L > X_C$ and circuit behaves like an RL circuit. Hence the circuit is said to be more inductive than capacitive.

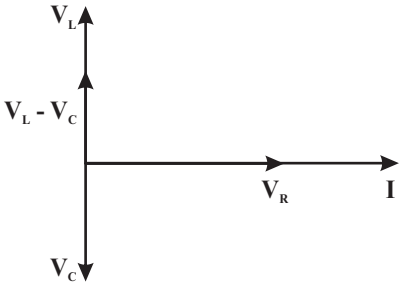


Figure 20.22
Phasor diagram of RLC series circuit

20.5.4 Power in A.C Circuits:

The calculation of power dissipation in a resistor involves the formula $P = VI$. However, this equation is not applicable in circuits where capacitors and inductors are connected to an alternating voltage source. The reason being that in a purely resistive circuit, the current and voltage are in phase. Whereas circuits containing capacitive or inductive elements exhibit a lead or lag of $\frac{\pi}{2}$ radians between the current and voltage, respectively, disrupting the applicability of this relationship. Therefore, in such a cases to determine power dissipation, project V onto the direction of referenece current phasor as shown in figure 20.18 (b). This approach ensures that the applied voltage is in phase with the current. Hence power dissipation in AC circuit is expressed as follows

$$P = VIcos\theta$$

where$cos\theta$ represents power factor.

Worked Example 20.6

A resistor (R) of 10 Ω and an inductor (L) of 0.05 H are connected in series. The AC voltage source has a peak voltage of 100 V and a frequency of 50 Hz.
Find:

1. The impedance of the circuit.
2. The RMS current through the circuit.
3. The average power consumed by the circuit

Solution:

Step 1: Write the known quantities and point out quantities to be found

R=10 Ω

L=0.05 H

V_{peak}= 100V

f= 50 Hz.

Step 1: Impedance of the Circuit (Z):

$$X_l = \omega_L = 2\pi fL = 2\pi \times 50 \times 0.05 = 15.7\ \Omega$$

$$Z = \sqrt{R^2 + X_L^2} = \sqrt{10^2 + 15.7^2} = \sqrt{100 + 246.49} = \sqrt{346.49} = 18.6\Omega$$

Step 2: RMS Current (I):

$$V_{RMS} = \frac{V_{peak}}{\sqrt{2}} = \frac{100}{\sqrt{2}} = 70.7V$$

$$I_{RMS} = \frac{V_{RMS}}{\sqrt{2}} = \frac{70.7}{18.6} = 3.8A$$

Step 3: Average Power (P):

$$P = V_{RMS} \times I_{RMS} \times \cos\phi$$

$$\text{Where } \cos\phi = \frac{R}{Z} = \frac{10}{18.6} = 0.54$$

$$P = 70.7 \times 3.8 \times 0.54 = 145W$$

Self-Assessment Questions:

1. Compare and contrast the flow of AC through resistors versus capacitors in terms of phase shift.
2. Describe the conditions under which maximum power transfer occurs between a source and a load in an AC circuit.

20.5.5 Resonant Frequency:

For a certain frequency the capacitive and inductive reactance becomes equal, $X_C = X_L$. This frequency is called resonant frequency f_r and circuit is said to be in resonance state. To calculate the resonant frequency, use the fact that

$$X_C = X_L$$

$$\frac{1}{2\pi fc} = 2\pi fL \text{ or} f_r = \frac{1}{2\pi\sqrt{LC}} \ldots\ldots (20.19)$$

From figure 20.23 note that X_C *and* X_L are in opposite direction. Therefore, at resonant frequency they cancel each effect in circuit. Now opposition to the current flow is solely offered by

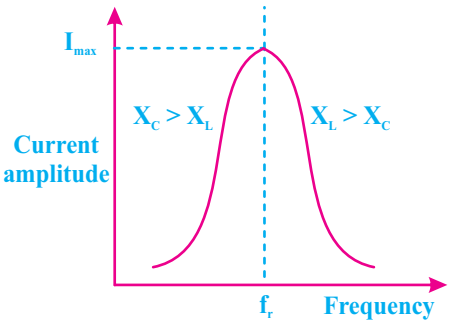


Figure 20.23 the resonant frequency

resistor, resulting in maximum current to flow through the circuit, moreover either side of resonant frequency the current in the circuit decreases, as shown in figure 20.23. RLC series circuit is an important circuit it is used in; Electronic filters, Electromagnetic Interference (EMI) Suppression, Tuned Circuits, etc

20.5.6 Parallel RL Circuit:

In some respects the circuit of figure 20.24 is similar to the purely inductive parallel circuit.

For instance, applied voltage V is still the quantity which is common to both components and is therefore plotted in standard position in the phasor diagram. Also the magnitude of the individual branch currents is determined by the opposition (reactance) of the individual branches. The figure 20.25 shows a composite diagram of waveforms and phasors. Since the phasor diagram shows that the two branch currents are not in phase, it will be necessary to use phasor addition in order to determine the total current.

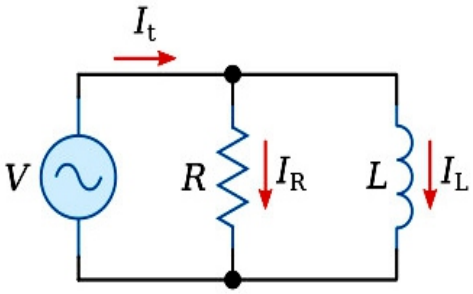


Figure 20.24 Parallel RL Circuit

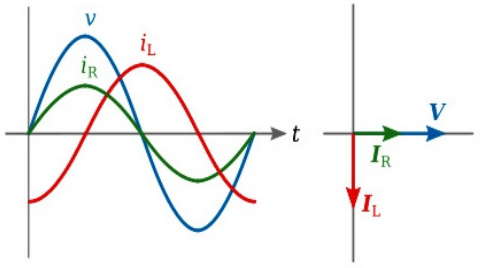


Figure 20.25 Waveforms and phasors

20.5.7 Parallel RC Circuit:

Parallel RC circuits may be resolved in much the same way as are parallel RL circuits. The figure 20.26 illustrates a parallel RC circuit.

The figure 20.27 shows a composite diagram of waveforms and phasors as per circuit conditions. The current phasors IR and IC are out of phase; therefore, phasor addition must be used to determine total current. The solving of an RC circuit follows the method previously applied to LR circuits.

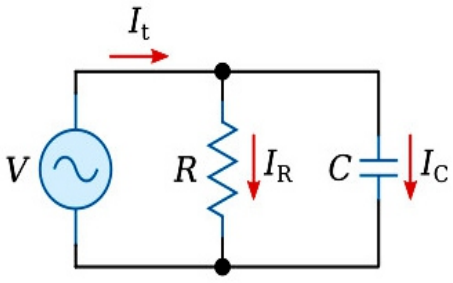


Figure 20.26 Parallel RC Circuit

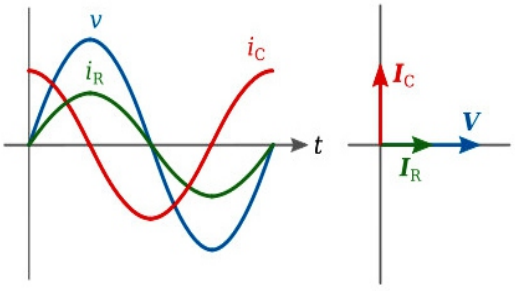


Figure 20.27 Waveforms and phasors

20.5.8 Parallel RLC AC Circuit:

A Parallel RLC AC Circuit is one where the resistor, inductor, and capacitor are connected in parallel to each other and the AC source. In this configuration, the voltage across each component is the same, but the currents through them differ. The parallel arrangement affects the overall impedance and current distribution in the circuit, making it distinct from its series counterpart. In a parallel circuit, the voltage V (RMS) across each of the three elements remains the same. Hence, for convenience, the voltage may be taken as a reference phasor.

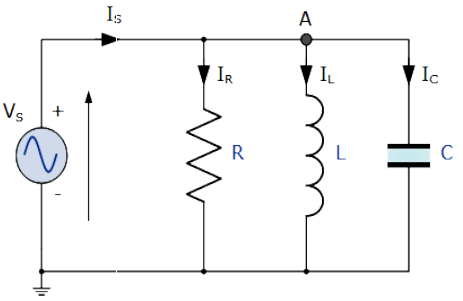


Figure 20.28 Parallel RLC AC Circuit

In the parallel RLC circuit, the supply voltage V_S is common to all three components, while the supply current I_S consists of three parts: the current through the resistor (I_R), the current through the inductor (I_L), and the current through the capacitor (I_C). The current flowing through each branch, and therefore through each component, will differ from one another and from the supply current I_S.

- The total current drawn from the supply is not simply the arithmetic sum of the three individual branch currents but their vector sum.
- Similar to the series RLC circuit, we can solve this circuit using the phasor or vector method. However, in this case, the vector diagram will use the voltage as its reference, with the three current vectors plotted relative to this reference voltage.
- The phasor diagram for an AC RLC parallel circuit is created by combining the individual phasors for each component and adding the currents vectorially.

Since the voltage across the circuit is common to all three circuit elements, we can use it as the reference vector, with the three current vectors drawn relative to this reference at their corresponding angles. The resulting vector current I_S is obtained by first adding the vectors I_L And I_C, and then adding this sum to the vector I_R. The angle between V and I_S.

20.5.9 Phasor Diagram of a Parallel RLC Circuit:

From the phasor diagram of AC RLC parallel circuit given below, we observe that the current vectors form a right triangle, with the hypotenuse represented by I_S, the horizontal axis by I_R, and the vertical axis by I_L–I_C. This configuration forms what is known as a Current Triangle. Consequently, we can apply Pythagoras's theorem to this current triangle to mathematically determine the individual magnitudes of the branch currents along the x-axis and y-axis. This will allow us to calculate the total supply current I_S of these components, as illustrated in figure. The circuit's phase angle is give.

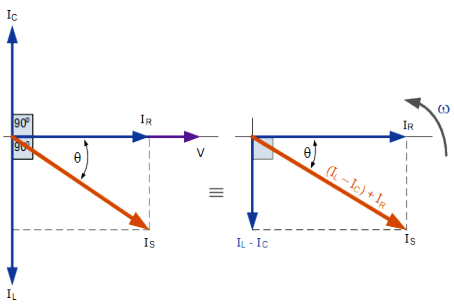


Figure 20.29 Phasor Diagram of a Parallel RLC Circuit

20.5.10 Resonance of Parallel RLC AC Circuit:

A parallel circuit containing a resistance, R, an inductance, L and a capacitance, C will produce a parallel resonance (also called anti-resonance) circuit when the resultant current through the parallel combination is in phase with the supply voltage. At resonance there will be a large circulating current between the inductor and the capacitor due to the energy of the oscillations, then parallel circuits produce current resonance.

A parallel resonant circuit stores the circuit energy in the magnetic field of the inductor and the electric field of the capacitor. This energy is constantly being transferred back and forth between the inductor and the capacitor which results in zero current and energy being drawn from the supply.

In the solution of AC parallel resonance circuits we know that the supply voltage is common for all branches, so this can be taken as our reference vector. Each parallel branch must be treated separately as with series circuits so that the total supply current taken by the parallel circuit is the vector addition of the individual branch currents.

$$I_R = \frac{V}{R}$$

$$I_L = \frac{V}{X_L} = \frac{V}{2\pi fL}$$

$$I_C = \frac{V}{X_C} = V.2\pi fC$$

Therefore, I_T = vector sum of $(I_R + I_L + I_C)$

$$I_T = \sqrt{{I_R}^2 + (I_L + I_C)^2}$$

At resonance, currents IL and IC are equal and cancelling giving a net reactive current equal to zero. Then at resonance the above equation becomes.

$$I_T = \sqrt{{I_R}^2 + (0)^2} = I_R$$

Therefore, the circuit current at this frequency will be at its minimum value of V/R and the graph of current against frequency for a parallel resonance circuit is shown in figure.

The frequency response curve of a parallel resonance circuit shows that the magnitude of the current is a function of frequency and plotting this onto a graph shows us that the response starts at its maximum value, reaches its minimum value at the resonance frequency when IMIN = IR and then increases again to maximum as f becomes infinite.

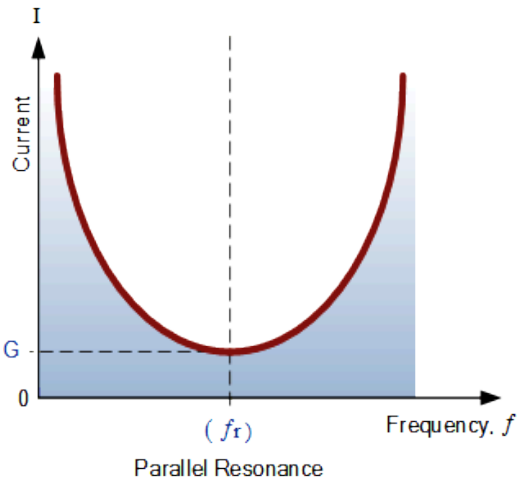


Figure 20.30 The graph of current against the frequency of parallel resonance circuit

The result of this is that the magnitude of the current flowing through the inductor, L and the capacitor, C tank circuit can become many times larger than the supply current, even at resonance but as they are equal and at opposition (180^{o} out-of-phase) they effectively cancel each other out.

As a parallel resonance circuit only functions on resonant frequency, this type of circuit is also known as an **Rejecter Circuit** because at resonance, the impedance of the circuit is at its maximum thereby suppressing or rejecting the current whose frequency is equal to its resonant frequency. The effect of resonance in a parallel circuit is also called "current resonance".

Self-Assessment Questions:

1. What factors determine the resonant frequency of an LC circuit?
2. How does the quality factor (Q-factor) relate to the sharpness of resonance in a circuit?

Worked Example 20.7

An inductor in for of solenoid have reactance of 500Ω is connected with alternating voltage 220V and frequency of 50 Hz. Calculate the inductive reactance of the solenoid.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$V = 220V,$
$f = 50\ Hz,$
$X_L = 500\Omega$
$Find\ L = ?$

Step 2: Write the formula and rearrange if necessary

$$X_L = 2\pi f L$$

Step 3: Substituting the give values in formula, we get

$$L = \frac{X_L}{2\pi f} = \frac{500}{6.28 \times 50} = 1.59 Henry$$

20.5.11 Maximum Power Transfer Theorem:

According to this theorem, maximum power is transferred from a source to a load when the impedance of the source matches with the impedance of the load.

In previous sections we have discussed that impedance is a measure of the opposition a circuit presents to the flow of alternating current (AC). Impedance includes both resistance and reactance, where reactance is the opposition due to capacitance or inductance.

When the impedance of the source matches the impedance of the load, the conditions for maximum power transfer are met. Under these circumstances, the electrical energy is efficiently transferred from the source to the load, resulting in the maximum power transfer.

Mathematically, suppose R_L denotes the resistance of load and R_s is source resistance.

Since $R_L = R_s$ as shown in figure 20.31 so maximum power transfer will occur.

In practical terms, achieving maximum power transfer is often important in designing electronic circuits to optimize efficiency and minimize power loss.

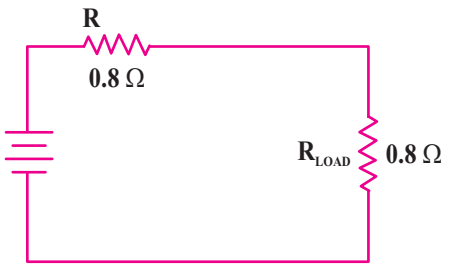


Figure 20.31 maximum power transfer

Worked Example 20.8

For an RLC circuit in parallel with a 5 Ω resistor, a 0.02 H inductor, and a 0.005 F capacitor, calculate the resonant frequency f_r?

Solution:

Step 1: Write the known quantities and point out quantities to be found

$L=0.02H$

$C=0.005F$

$f_r = ?$

Step 2: Write the formula and rearrange if necessary

$$f_r = \frac{1}{2\pi\sqrt{LC}}$$

Step 3: Substituting the give values in formula, we get

$$f_r = \frac{1}{2\pi\sqrt{0.02 \times 0.005}}$$

$$f_r = 15.92\ Hz$$

20.6.1 Metal Detectors:

The oscillator circuit is used in metal detectors. Metal detectors used for security checks operate on the principle of electromagnetic induction. A simplified explanation of working of metal detector is described below:

1. **Generating an Electromagnetic Field:**
   The metal detector contains a coil of wire through which an electric current flow, creating a magnetic field around the coil. This coil is often housed in a special arrangement, such as a loop or wand.
2. **Interaction with Metals:**
   When a conductive metal object is brought into the vicinity of the electromagnetic field, it disturbs the field. This disturbance induces a secondary magnetic field in the metal object.

3. **Eddy Currents:**
   The changing magnetic field induces circulating electric currents within the metal object, known as eddy currents. These eddy currents, in turn, generate their own magnetic fields.
4. **Detection of Changes:**
   The metal detector has a receiver coil or coils that are in close proximity to the transmitter coil. The receiver coil(s) detect changes in the magnetic field caused by the presence of the metal.
5. **Alert Mechanism:**
   When the metal detector senses a significant change in the magnetic field, indicating the presence of a metal object, it triggers an alert. This alert can be in the form of an audible sound, a visual signal, or both, depending on the design of the metal detector.

20.6.2 The Electrocardiograph:

The electrocardiograph (ECG) is a medical diagnostic tool used to record the electrical activity of the heart over a period of time.

Principle:

The principle behind the electrocardiograph is based on the electrical signals generated by the heart during each heartbeat.

Working:

The heart is a muscular organ that contracts rhythmically to pump blood throughout the body. This contraction is initiated and coordinated by electrical signals generated within the heart. Electrodes are attached to specific points on the skin of the patient. The standard placement involves attaching electrodes to the limbs and chest. These electrodes are conductive and are used to detect the electrical signals produced by the heart. The electrical signals produced by the heart are picked up by the electrodes. The detected electrical signals are amplified to make them more measurable and are then recorded on a graph or displayed on a monitor. The resulting graph is called an electrocardiogram. The ECG graph represents the electrical activity of the heart over time. It consists of waves and intervals, each of which corresponds to a specific phase of the cardiac cycle. Physicians analyze the ECG to gather information about the heart's rhythm, rate, and various other aspects of its electrical activity. Deviations from the normal ECG pattern can indicate cardiac abnormalities, such as arrhythmias, ischemia, or other heart conditions.

20.6.3 Oscillator Circuit:

An oscillator circuit is an electronic circuit that generates a continuous periodic signal at a specific frequency.

Oscillator circuits produce the carrier wave, which serves as the central frequency around which the information-carrying signal is modulated.

Oscillator circuits are designed to be tunable, allowing broadcasters to set the carrier frequency to a specific value. This turnability is essential for assigning unique frequencies to different radio stations, preventing interference between them. Oscillators are designed to be

modulated by an information-carrying signal. In amplitude modulation, the amplitude of the carrier wave is varied according to the audio signal, while in frequency modulation; the frequency of the carrier wave is modulated. This modulation process allows the transmission of audio information. The carrier wave itself does not convey information; it serves as a medium to carry the modulated information signal. The oscillator generates a stable carrier wave, ensuring that the modulated signal can be reliably transmitted over long distances. The carrier wave, when modulated with the audio signal, forms the composite radio wave that propagates through space. This transmitted signal can be received by radio receivers tuned to the carrier frequency.

In summary, the oscillator circuit is the heartbeat of a radio transmitter, generating the carrier wave that facilitates the transmission of modulated information signals. Its stability, tunability, and modulation capabilities are pivotal for effective and interference-free radio broadcasting.

20.6.4 Resonance in Tuning circuit of Radio:

In radio tuning circuits resonance is a important phenomenon that enables the selective reception of a desired radio frequency while rejecting others.

Principle:

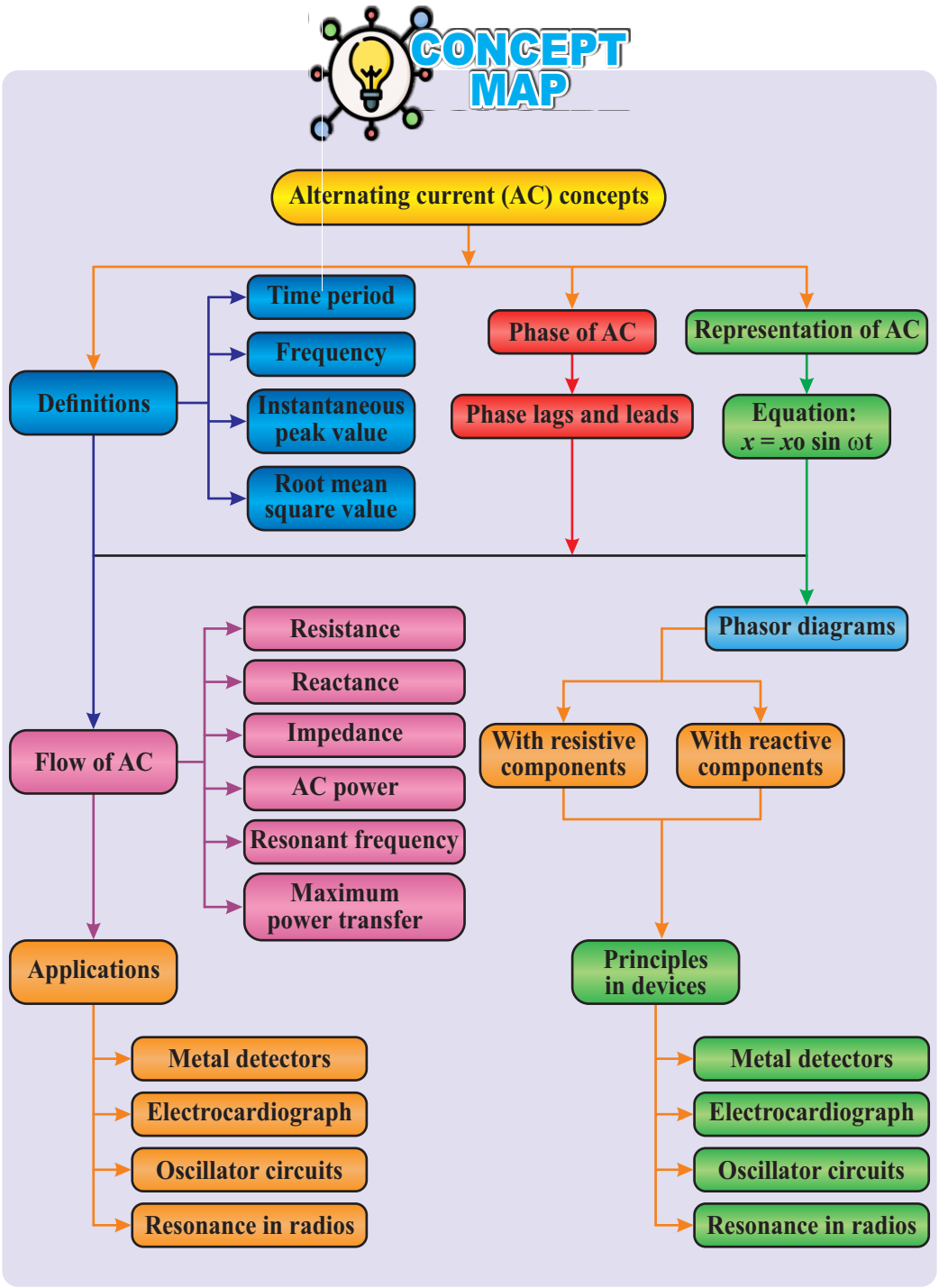
The principle of resonance is crucial in the tuning circuits of radio to enhance the reception of specific radio frequencies. Resonance occurs when the inductive and capacitive reactance in a circuit cancel each other out, resulting in a condition where the circuit efficiently absorbs energy at a particular frequency. Resonance is integral to the functioning of radio tuning circuits. It ensures that radios can selectively receive signals from desired stations with clarity and efficiency, making it possible to enjoy various broadcasted content without interference from other frequencies.

Importance of Broadcasting:

Broadcasting remains a powerful medium with far-reaching impacts on society. It informs, educates, entertains, and connects people, playing a crucial role in cultural preservation, economic development, and social cohesion. Its ability to reach a wide audience makes it an indispensable tool for communication in the modern world.



- ✓ AC or alternating current refers to the electric charge flow that periodically reverses direction, typically sinusoidal.
- ✓ The frequency of AC current, determines how many times per second the AC current reverses direction.
- ✓ Root Mean Square (RMS equivalent steady DC voltage producing the same power in a resistive load as the AC voltage.
- ✓ Current and Voltage are in phase in a resistive load known as Ohm's Law for AC current as $I = \frac{V_{RMS}}{R}$
- ✓ In pure capacitive circuit the current leads the voltage by 90 degrees or $\frac{\pi}{2}$ radians
- ✓ In inductive circuit the current lags the voltage by 90 degrees or $\frac{\pi}{2}$ radians
- ✓ The combined effect of resistance and reactance in RC or RLC circuit is known as impedance
- ✓ Resonant frequency is the frequency at which an oscillating system naturally vibrates with the greatest amplitude. At resonant frequency the impedance of the system is at its minimum.
- ✓ The RLC series circuit configuration results in high impedance at resonance and a lower overall current compared to the individual component currents.
- ✓ The RLC parallel circuit configuration the components are connected in parallel branches, providing multiple current paths. At resonance, the impedance is at its minimum, allowing a higher overall current flow compared to the series circuit.
- ✓ Maximum power is transferred from a source to a load when the load impedance is equal to the source impedance.
- ✓ ECG is a medical device that records the electrical activity of the heart over time.
- ✓ Oscillator Circuit is an electronic circuit that produces a periodic, oscillating signal, often a sine wave or square wave.
- ✓ Choke Coil is an inductor used to block higher-frequency AC signals while allowing lower-frequency or DC signals to pass.
- ✓ Resonance is the condition in a circuit when the inductive reactance and capacitive reactance are equal in magnitude but opposite in phase, resulting in a maximum voltage or current at a particular frequency.


CONCEPT MAP
Alternating current (AC) concepts
Definitions
Time period
Frequency
Instantaneous peak value
Root mean square value
Phase of AC
Phase lags and leads
Representation of AC
Equation: x = xo sin ωt
Phasor diagrams
With resistive components
With reactive components
Flow of AC
Resistance
Reactance
Impedance
AC power
Resonant frequency
Maximum power transfer
Applications
Metal detectors
Electrocardiograph
Oscillator circuits
Resonance in radios
Principles in devices
Metal detectors
Electrocardiograph
Oscillator circuits
Resonance in radios

EXERCISE

Section (A): **Multiple Choice Questions (MCQs)**

Choose the correct answer:

1. In alternating current, the flow of electric charge periodically reverses direction. The term for this specific phenomenon is:
   (a) Alternation (b) Oscillation (c) Cycle (d) Resistance
2. The frequency of standard household alternating current in Pakistan is:
   (a) 50 Hz (b) 60 Hz (c) 100 Hz (d) 120 Hz
3. The parameter for measuring the rate of change of alternating current or voltage with respect to time is:
   (a) Amplitude (b) Frequency (c) Phase angle (d) Peak Voltage
4. In AC circuits, what is the phase difference between current and voltage in a purely resistive circuit?
   (a) 0^0 (b) 45^0 (c) 90^0 (d) 180^0
5. A device used to measure the root mean square (RMS) value of an AC voltage or current is:
   (a) Voltmeter (b) Oscilloscope (b) Ammeter (d) AVO meter
6. In an AC circuit. The term "reactance" represent the:
   (a) Ohmic resistance only
   (b) Resistance to DC currents only
   (c) Capacitive and inductive opposition to current flow
   (d) Total impedance, including resistance and phase angle
7. A type of a circuit characterized by equal values of resistance and reactance is:
   (a) Resistive (b) Inductive (c) Capacitive (d) Resonant
8. In an AC sine wave, the peak-to-peak voltage is:
   (a) Double the peak voltage
   (b) The difference between the peak voltage and zero
   (c) The sum of the peak voltage and zero
   (d) The difference between the positive and negative peak voltages
9. What is the power factor of a purely resistive circuit?
   (a) 0 (b) 0.5 (c) 1 (d) -1
10. The purpose of a transformer in an AC power distribution system.
    (a) To convert AC to DC (b) To step up or step down voltage
    (c) To store electrical energy (d) To regulate current flow

Section (B): **CRQs (Short Answered Questions):**

1. Define RMS voltage and explain its significance in AC circuits.
2. Explain the difference between peak voltage, RMS voltage, and average voltage in AC circuits.
3. Define alternating current (AC) and explain how it differs from direct current (DC).
4. Define reactance in AC circuits and differentiate between capacitive and inductive reactance.

5. Describe the behavior of capacitors and inductors in AC circuits.
6. Explain the phase relationship between voltage and current in capacitive and inductive AC loads.
7. What is resonance in AC circuits? Discuss its conditions and applications.
8. Discuss the role of transformers in AC circuits and explain how they work.
9. Explain the behavior of an RLC series circuit in terms of impedance, resonance, and phase angle.
10. Describe the operation and applications of a transformer in AC circuits.

Section (C): ERQs (Long Answered Questions):

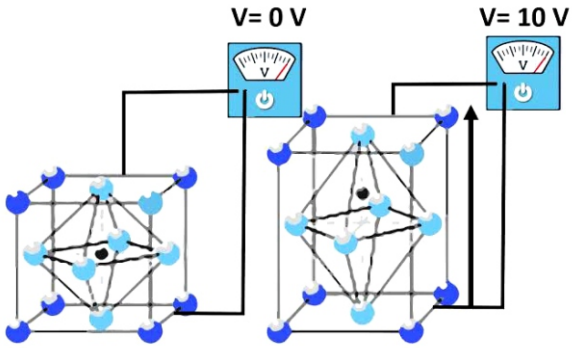
1. Explain the concept of phasor in the context of AC voltage. How are phasors used to represent sinusoidal voltages?
2. Describe the concept of impedance in AC circuits. How does impedance differ from resistance, and what are the units of impedance?
3. How does the reactance of an inductor and a capacitor change with frequency in an AC circuit? Provide an explanation based on the fundamental formulas?
4. Discuss the concept of resonance in RLC circuits. What conditions lead to resonance, and how does it affect the behavior of the circuit?
5. Compare and contrast series and parallel resonance in RLC circuits. What are the key differences between the two resonance configurations?
6. Explain the transient response of an RLC circuit when initially connected to an AC source. What happens to the currents and voltages in the circuit?
7. Why alternating current (AC) is commonly used for long-distance power transmission?
8. Describe the purpose and function of a choke coil in AC circuits.

Section (D): Numerical:

1. A resistor (R) of 20 ohms is connected in series with a capacitor (C) of 10 μF in an AC circuit with a frequency of 50 Hz. Calculate the total impedance? **(3182 ohms)**
2. For an inductor with an inductance (L) of 0.5 H and a frequency of 100 Hz, calculate the inductive reactance? **(314.16 ohms)**
3. In an RL circuit, the resistance (R) is 30 ohms, and the inductance (L) is 0.2 H. Calculate the total impedance at a frequency of 60 Hz? **(81.1ohms)**
4. In an RC circuit, the resistance (R) is 50 ohms, and the capacitance (C) is 20 μF. Calculate the capacitive reactance? **(31.83 ohms)**
5. An AC circuit has a resistance of 40 ohms, an inductive reactance of 30 ohms, and a capacitive reactance of 20 ohms. Draw the impedance triangle and calculate the total impedance? **(41.2ohms)**
6. In a series RL circuit, the resistance (R) is 25 ohms, and the inductance (L) is 0.1 H. Calculate the phase angle and impedance at a frequency of 80 Hz? **(63.8^0, 502 ohms)**
7. In a parallel RC circuit, the resistance (R) is 60 ohms, and the capacitance (C) is 30 μF. Calculate the total current flowing through the circuit at a frequency of 120 Hz? **(60 V)**
8. In an RLC circuit, the resistance (R) is 50 ohms, the inductance (L) is 0.1 H, and the capacitance (C) is 50 μF. Calculate the resonance frequency? **(70.7Hz)**

Unit 21 Physics of Solids

Teaching Periods 11 **Weightage %** 07



The Piezoelectric effect is the ability of certain materials to generate an electric charge in response to mechanical stress, such as pressure, vibration, or stretching, causing the material's internal electric dipoles to align and produce a voltage across the material, with the magnitude of the charge being directly proportional to the applied stress, allowing the material to convert mechanical energy into electrical energy.

In this unit student should be able to:

- Distinguish between the structure of crystalline, glassy, amorphous and polymeric solids.
- Describe that deformation in solids is caused by a force and that in one dimension; the deformation can be tensile or compressive
- Define and use the terms Young's modulus, bulk modulus and shear modulus.
- Demonstrate the force-extension graphs for typical ductile, brittle and polymeric materials.
- Become familiar ultimate tensile stress, elastic deformation and plastic deformation of a material.
- Describe the idea about energy bands in solids.
- Classify insulators, conductors, and semiconductors on the basis of energy bands.
- Become familiar with the behavior of superconductors and their potential uses.
- Describe the applications of superconductors in magnetic resonance imaging (MRI), magnetic levitation trains, powerful but small electric motors and faster computer chips.
- Distinguish between dia, para and Ferro magnetic materials.
- Describe the concepts of magnetic domains in a material.
- Explain the Curie point. Classify hard and soft ferromagnetic substances.
- Describe hysteresis loss.
- Synthesis from hysteresis loop how magnetic field strength varies with magnetizing current.
- Identify the importance of hysteresis loop to select materials for their use to make them temporary magnets or permanent magnets.

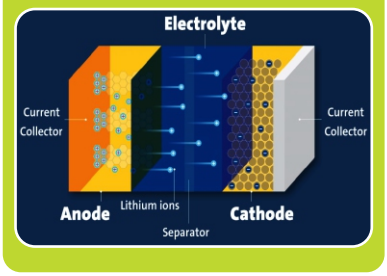
Introduction:

The Physics of solids is a branch of physics that deals with the study of the physical properties of solid materials, including their electronic, magnetic, optical, thermal and mechanical properties. It aims to understand the behavior of solid materials under different conditions, such as temperature, pressure and electromagnetic fields, using principles of quantum mechanics.It also helps to explain phenomenon like crystal structures, phase transitions, defects in material, and the interactions between electrons, ions and atoms in solids.

For example, a key hurdle to enhance solar power which is an affordable energy source. It is increasing the efficiency of photovoltaic cells that convert solar radiation to electricity, making solid materials in solar cells as conductive as possible

DO YOU KNOW?

John B, who credited for the identification and initial development of the first lithium-ion (Li-ion) batteries. He realized how a battery with lithium anode could provide high charge density and thus give us the pre-requisite technology for electric cars, and smart phones.



21.1 Classification of Solids:

Solids are usually crystalline substances their molecules are arranged in a definite pattern and in fixed positions.The way in which a solid behaves depends on its internal structure and there are three types of solid.

(i) Crystalline solids (ii) Amorphous solids (iii) Polymeric solids

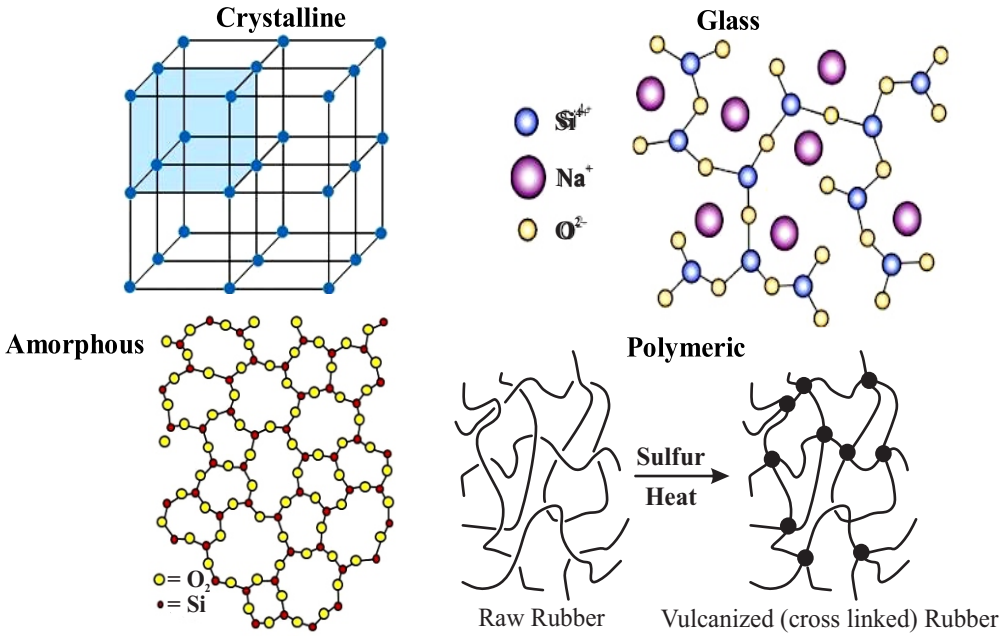


Figure 21.1 structures of crystalline, glassy, amorphous and polymeric solids

21.1.1 Distinguish between structure of crystalline, glassy, amorphous and polymeric solids:

Property	Crystalline Solids	Glassy Solids	Amorphous Solids	Polymeric Solids
Atomic Arrangement	Regular, Repeating pattern.	Random, short-range order.	Random, no long-range order.	Long chains or networks of repeating units.
Melting Point	Sharp, Specific temperature.	Gradual, over a range of temperature.	Gradual, over a range of temperatures.	Varies, typically lower than crystalline solids.
Rigidity	Very rigid and well-defined structure.	Rigid, but less than crystalline solids.	Not as rigid as crystalline solids.	Can be flexible or rigid.
Transparency	Can be transparent or translucent.	Transparent or opaque depending on composition.	Opaque.	Depending on structures, can be transparent or opaque.
Examples	Diamond, salt crystals, silicon.	Window glass, certain plastics, amorphous metals.	Rubber, some plastics, glass.	Polyethylene, PVC, nylon.

21.1.1 Deformation:

In material science, deformations refer to modifications of the shape or size of an objectdue to applied forces or a change in temperature. Deformation is usually caused by forces such as tensile (pulling) or Compressive (pushing) as shown in figure 21.2. As deformations occurs, internal inter-molecular forces arise that oppose the applied force.Different types of deformations may result from variations in type of material, size and the forces applied.

Deformation in one dimension (1D):

Stress:

It is usually defined as force applied to a material per unit area.Stress is a quantity which defines the magnitude of force that cause deformation.

Strain:

Strain is a measure of how much a material deforms (stretches or compresses) when a force is applied to it. It is calculated by dividing the change in length by the original length of the material:

When pulling force on an object is applied which causes elongation, like the stretching of an elastic band, we call such stress a tensile stress. When forces cause a compression of an object, we call it a compressive stress. As illustraded in Figure 21.2.

The greater the stress, the greater the strain; however, the relation between strain and stress does not need to be linear. When stress is sufficiently low, deformation is causesd in direct proportion to the stress value. The proportionality constant in this relation is calledthe **elastic modulus**. In the linear limit of low stress values, the general relation between stress and strain is known as Hooke's law is given as under

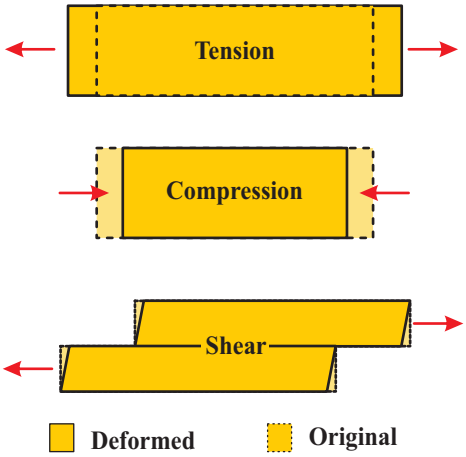


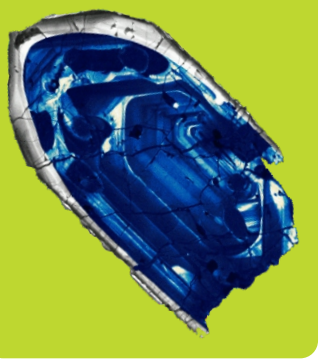
Figure 21.2 Deformation in one dimension

Stress =(elastic modulus)×Strain

We can also see from above equation that when an object is characterized by a large value of elastic modulus, the effect of stress is small. On the other hand, a small elastic modulus means that stress produces large strain and noticeable deformation happens. For example, a stress on a rubber band produces larger strain (deformation) than the same stress on a steel band of the same dimensions because the elastic modulus for rubber is two orders of magnitude smaller than the elastic modulus for steel.

DO YOU KNOW?

The zircons, the oldest known terrestrial materials, have helped portray how the Earth's crust formed during the first geologic eon of the planet. "This may also help us understand how other habitable planets would form.



21.2 Mechanical Properties of Solids

Mechanical properties of solids refer to the characteristics that describe how a solid material responds to applied forces or deformation; some important mechanical properties of solid are Elasticity, Ductility, Brittleness, and Hardness etc.

21.2.1 Young's Modulus (modulus of elasticity)

It is a mechanical property of solids that measures the stiffness or elasticity of a solid material and it computes how much a material will deform (stretch or compress) under a given force.

It is denoted by the symbol Y and has units of $Y = \frac{\text{Force}}{\text{Area}}$ or (Pascal, Pa or N/m^2).

It is a measure of a material's resistance to elastic deformation when subjected to an external force. The higher the Young's Modulus, the stiffer the material.

For example, steel has a higher Young's Modulus than rubber, which makes steel more resistant to deformation than rubber when subjected to a force.

The amount of extension of an object depends upon

i) Applied force
ii) Material
iii) Dimension

If we compare rods made of the same material having different lengths and cross-sectional areas, it is found that the longer the object, the more it elongates for a given stress; and the thicker it is, the less it elongates because strain is proportional to the original length and inversely proportional to the cross-sectional area.

Then Hooke's law can be written for Young's Modulus is defined as the ratio of stress to strain.

$$Y = \frac{Stress}{Strain}$$

$$Y = \frac{\frac{F}{A}}{\frac{\Delta L}{L}}$$

$$Y = \frac{F \times L}{\Delta L \times A} \ldots\ldots\ldots (21.1)$$

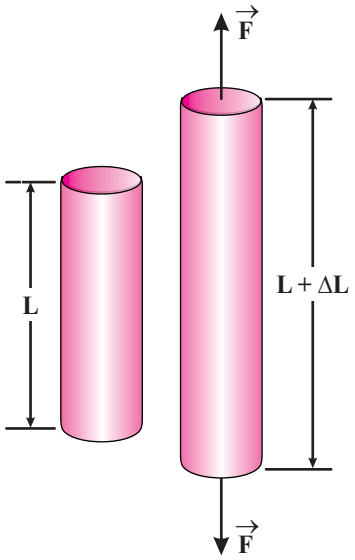


Figure 21.3 longitudinal stress

Worked Example 21.1

Calculate the change in length of the upper leg bone (the femur) when a 70.0 kg man supports 62.0 kg of his mass on it, assuming the bone to be equivalent to a uniform rod that is 40.0 cm long and2.0 cm in radius. Young's modulus for bones is 19×10^{9} N/m^2.

Solution:

Step 1: data

Mass of man = 70.0 kg
Mass supported by leg = 62.0 kg
Length of leg, L = 40.0 cm = 0.4 m
Radius of bone = 2.0 cm = 0.02 m
Young's modulus, Y = 19×10^{9} N/m^2.
Change in length, ΔL = ?

Step 2:

F= W = mg = 62.0 × 9.8 = 607.6 N

Step 3: Calculate the cross-sectional area:

$A = \pi r^2 = (3.14)(0.02)^2 = 12.56 \times 10^{-4} m^2$

Now calculate the change in length

$$Y = \frac{F \times L}{\Delta L \times A}$$

$$\Delta L = \frac{F\ x\ L}{A\ x\ Y}$$

$$= \frac{607.6 \times 0.40}{19 \times 10^{9} \times 1.256 \times 10^{-4}}$$

$$\Delta L = 1.01 \times 10^{-5}\ m$$

Result: $\Delta L = 1.01 \times 10^{-5}\ m$

Self-Assessment Questions:

1. Calculate the amount of stretch in steel cable. Suspension cables are used to carry gondolas at Murree Hills. Consider a suspension cable that contains an unsupported span of 1.5 km with diameter of 3.0 cm, with maximum tension it can withstand is 3.0×106N.

21.2.2 Shear Modulus:

Shear modulus, also known as the modulus of rigidity is a mechanical property that describes a material's resistance to shear deformation when a force is applied perpendicular to the material's surface. It is denoted by the symbol G or S and has units of force per unit area, typically expressed in Pascal (Pa) or Giga Pascal (GPa).

Shear modulus is the measure of the rigidity of the body which is the ratio of shear stress to shear strain.

Consider a rigid body as shown in figure 21.4. When acted upon by tangential force to twist it.

The shear stress is F/A and shear strain is ΔX/L. Therefore, the shear modulus becomes

$$G = \frac{\frac{F}{A} \times L}{\Delta x}$$

Table: No 21.1

Material	G(10^9Nm^{2}or GPa)
Aluminum	25
Brass	36
Copper	42
Glass	23
Iron	70
Lead	5.6
Nickel	77
Steel	84
Tungsten	150
Wood	10
Diamond	450
Rubber	0.1

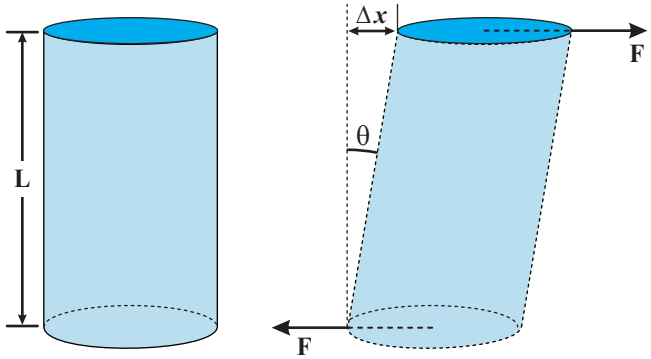


Figure 21.4: shear modulus

$$ShearingStrain = \Delta X = \frac{tan\theta}{l} \quad \ldots\ldots\ldots (21.2)$$

Worked Example 21.

The area of the upper face of a rectangular block is 0.25 m^2 and the lower face is fixed. The height of the block is 1.5 cm, a shearing force applied to the top face which produces a displacement of 0.0152 mm. Find the strain, stress and the shearing force. Modulus of rigidity is 4.5×10^{10} N/m^2.

Solution:

Step 1: **data**

$Area\ under\ shear\ = 0.25\ m^2$

$Height\ of\ block\ = 1.5\ cm\ = 1.5\times10^{-2}\ m$

$Displacement\ of\ top\ face\ = 0.0152\ mm\ = 0.0152\times10^{-3}\ m$

$Modulus\ of\ rigidity\ = G\ = 4.5\times10^{10}\ N/m^2.$

$Shear\ Strain$ =?

$Shear\ Stress$ =?

Shearing Force =?

Step 2: **Formula:**

$Shearing\ Strain\ = \Delta X/L$

Step 3: **Calculations: Shearing Strain**

$$Shearing\ Strain = \frac{0.0152\times10^{-3}}{1.5\times10^{-2}}$$

Shearing Strain = 1.01×10^{-3}

Shear Stress = G × Shearing Strain

Shear stress= $4.5\times10^{10}\times 1.01\times10^{-3}$

Shearing Stress = 4.56×10^7 N/m^2

Shear Stress = F /A

F = Shear Stress × A

F = $4.56\times 10^7\times 0.25$

F = 1.14×10^7 N

Result: F = 1.14×10^7 N

21.2.3 Bulk Modulus:

Bulk modulus, also known as the modulus of compressibility, is a mechanical property that describes the resistance of a material to compression under uniform external pressure. It is denoted by the symbol K and has units of pressure (usually expressed in Pascal, Pa or Giga Pascal (GPa).

Bulk modulus is defined as the ratio of the change in pressure applied to a material to the resulting fractional change in volume that occurs in the material. Mathematically, it is expressed as:

$$\frac{\Delta V}{V} = -\frac{\Delta P}{B}$$

DO YOU KNOW?

Compressibility

The reciprocal of bulk modulus of elasticity is called as compressibility

$Compressibility = 1/B$

Its SI unit is m2 /N or Pa.

Material	Bulk modulus
Air	100KPa
Water	2.2 GPa
Steel	150GPa
Diamond	443 GPa

$$B = -\frac{\Delta P}{\frac{\Delta V}{L}} \quad \ldots\ldots\ldots\ldots\ldots (21.3)$$

where B is the bulk modulus, ΔP is the change in pressure applied to the material, ΔV is the resulting change in volume, and V is the original volume of the material.

21.2.4 Force– Extension Graph:

In material science, the force versus extension graph for a material gives the relationship between stress and strain. This is obtained by gradually applying a load to a test component and measuring its deformation according to tensile standards. These curves reveal many of properties of materials, such as the young's modulus, the yield strength, the ultimate tensile strength and so on. For different types of materials as shown in figure 21.5,

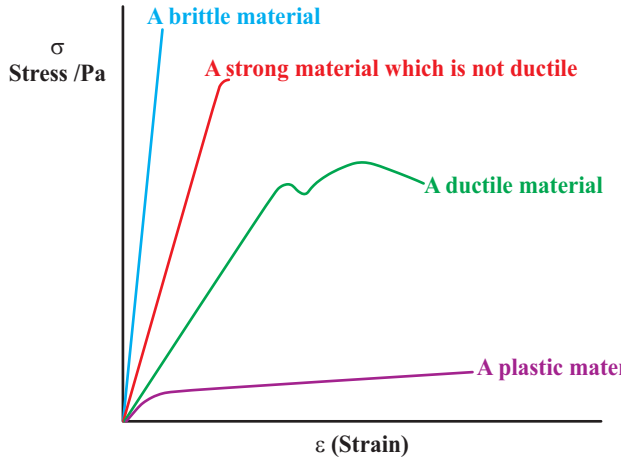


Figure 21.5 force extension graph reveals the properties of different materials

The blue line (brittle) represents the behavior of a typical elastic material, which follows Hooke's Law. As the material is stretched, the force applied to it increases linearly with the extension, up to a point called the yield point. The material eventually reaches a maximum point of extension, is called the ultimate extension or fracture point, beyond which it breaks or fractures.

The green line (ductile) represents the behavior of a typical viscoelastic material, which exhibits time-dependent deformation under a constant force. In this case, the material initially undergoes elastic deformation, similar to the elastic material, but then continues to deform slowly over time under a constant force, eventually reaching a maximum extension or creep limit.

DO YOU KNOW?

Viscoelastic materials combine the properties of rubber and slush, deforming and flowing when stressed. They absorb and dissipate energy over time, making them ideal for shock absorbers, vibration reduction, and noise dampening. Examples include rubber, plastics, oil, and molten glass, which are valuable in engineering for designing durable products like gaskets and automotive parts that withstand repeated stress and deformation.

DO YOU KNOW?

Silk is a natural polymer that is widely used worldwide in textiles. Silk threads are extremely light, but are very strong relative to their weight. It is extremely ductile, so it can bend and stretch without breaking.

The purple line (Plastic) represents the behavior of a typical plastic material, which exhibits permanent deformation when subjected to a force. In this case, the material undergoes plastic deformation immediately, with no linear region or yield point, and continues to deform permanently as the force is increased, until it ultimately fractures or breaks.

The Red line is showing strong material which is not ductile, like steel wires stretch very little and breaks suddenly.

Polymeric Materials:

Polymer curves are produced by stretching a sample at a constant rate through the application of a tensile force. By using a constant rate of testing the strain-rate dependency of polymer behavior is not allowed to dominate. However, it should be appreciated that polymers have a marked inherent time - dependence in their response to deformation, which sets their behavior apart from other classes' materials.

21.2.5 Graph of Ultimate tensile stress, elastic deformation and plastic deformation:

Figure 21.6, represents the graph of applied force versus elongation up to a point is called the proportional limit. Hooke's law is the best approximation for many common materials, and the curve is a straight line. Beyond this point, the graph deviates from a straight line, and no simple relationship exists between F and Δl. Nonetheless, up to a point farther along the curve called the elastic limit, the object will return to its original length if the applied force is removed. The region from the origin to the elastic limit is called elastic region. If the object is stretched beyond the elastic limit, it enters the plastic region: it does not return to the original length upon removal of the external force, but remains permanently deformed (such as bent paper clip shown in figure 21.7) such is elastic deformation. The maximum force that can be applied without breaking is called ultimate tensile strength of the material. If the stretching of bonds happens but the atoms do not slip past each other, when the stress is sufficient to permanently deform the metal i.e., soft metal like silver, gold, it is called plastic deformation.

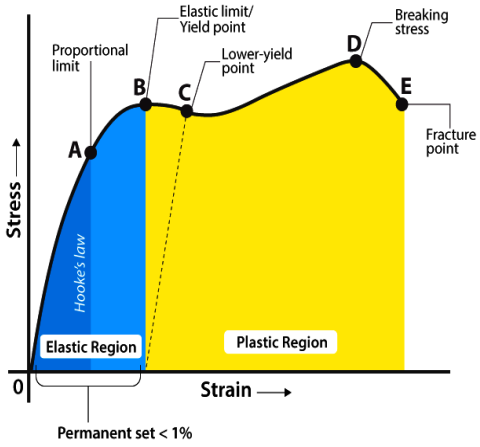


Figure 21.6 Applied force vs elongation



Figure 21.7 elastic deformation

21.3 Electrical Properties of Solids:

The electrical property of a good conductor depends on its ability to transmit energy without boiling, melting, or changing its composition in any way.

- **Resistivity:** The resistance per unit length and cross-sectional area is called resistivity. It varies from material to material. Its SI unit is Ohm-meter (Ω - m).
- **Conductivity:** It is a measure of a material's ability to conduct electric current, which is influenced by the availability and mobility of charge carriers within the solid.
- **Temperature Coefficient of resistivity:** The temperature coefficient of resistance is defined as the change in electrical resistance of a substance with respect to per degree rise in temperature. When the temperature increases, the process of electron collision becomes rapid and faster. As a result, the resistance will increase with the rise in temperature of the conductor.
- **The quantity of charge carriers' n,** the number of chare carriers per unit volume, can be found from measurement of the Hall Effect. Its SI unit is m^{-3}.

21.3.1 Energy Bands in Solids:

When isolated atoms comes together to form a solid, interactions between neighboring atoms cause the electron energy levels to split and overlap, creating continuous energy bands.

Valence Band:

The valence band is the range of energy levels occupied by electrons that are bound to atoms and participate in chemical bonding. It is the highest energy band that contains electrons under normal conditions.

Forbidden Band (Energy Gap):

The forbidden band, or energy gap, is the range of energy levels between the valence band and the conduction band where no electron states exist. This gap influences the material's electrical properties; a larger gap typically means the material is less conductive.

Conduction Band:

The conduction band is the range of energy levels where electrons are free to move throughout the material, allowing electrical conduction. Electrons in the conduction band are not bound to any particular atom and can carry electric current through the solid.

In the case of a diamond crystal, carbon atoms with the electronic structure $1s^2\ 2s^2\ 2p^2$ form two main energy bands as shown in figure, 21.8. The valence band and the conduction band are separated by an energy gap known as the "forbidden band." The valence band contains the lower

DO YOU KNOW?

The Pauli Exclusion principle states that "No two electrons in a given interacting system may have the same quantum state". Thus there must be a splitting of the discrete energy levels of the isolated atoms into new level belonging to the pair, other than to individual atoms

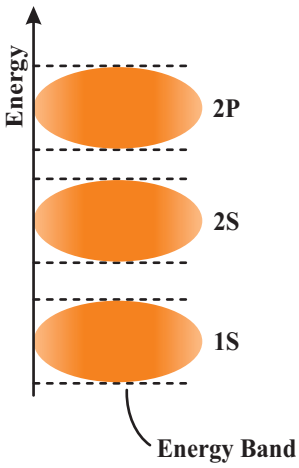


Figure 21.8 energy Bands

energy states, while the conduction band contains higher energy states. This energy gap prevents electrons in the valence band from easily moving to the conduction band, significantly influencing the electrical properties of the solid, such as its ability to conduct electricity.

21.3.2 Classification of Solids:

Depending on the electrical conductivity solids are classified into three main categories: conductors, semiconductors and insulators, as shown in figure 21.9.

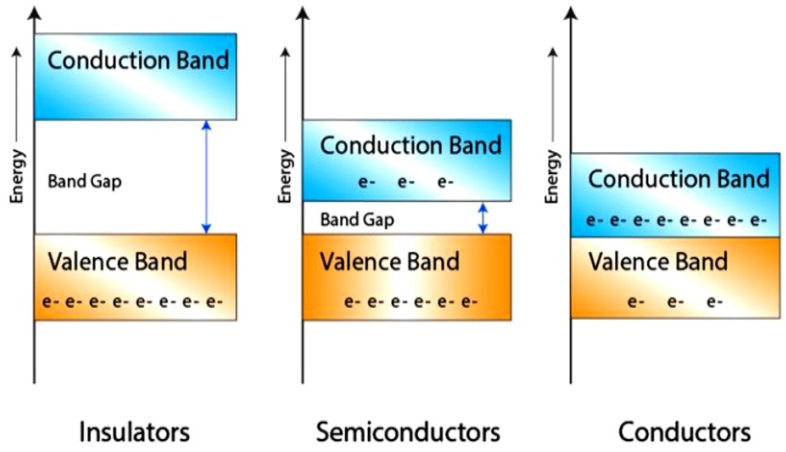


Figure 21.9 Classifications of Solids in three main categories: Insulators, Semiconductors and conductors

Insulators:

These materials do not conduct electricity. The band gap between the valence band and conduction band is very large. Even if a large amount of energy is provided to these solids, they do not conduct electricity. such as wood, plastics etc.

Conductor:

There is no gap between the conduction band and the valence band. Thus, electrons can easily flow from the valence band to the conduction band under the influence of an electric field, making them good conductors of electricity.

Semiconductors:

The gap between the conduction band and the valence band is very less; therefore, whenever sufficient energy is provided to the electrons in a semiconductor, electrons jump from the valence band to the conduction band. The conductivity of the semiconductors increases with an increase in temperature, for semiconductors, it lies in between 10^{-6} to 10^{-4} $(\Omega\ m)^{-1}$. There are two types of semiconductors:

a) Intrinsic semiconductor:

When energy is provided to intrinsic semiconductors like silicon and germanium, electrons leave their positions,

DO YOU KNOW?

Measurements show that the resistivity of super conductors is less than 4×10^{-25} Ω m, which is over 1016 times smaller than copper, and is considered to be zero in practice.

creating positive holes. Under an electric field, electrons and holes move in opposite directions, enabling electrical conductivity. Group IV elements are considered as an interinsic semiconductors

b) Extrinsic semiconductor:

Silicon and germanium have low conductivity at room temperature. Adding small amounts of group III or group V impurities, a process called doping, increases their conductivity, resulting in extrinsic semiconductors.

Intrinsic and extrinsic semiconductor are shown in Figure 21.10

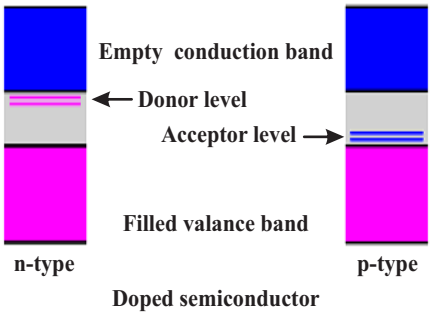


Figure 21.10
Intrinsic and extrinsic semi-conductor

21.3.3 Superconductor:

In 1911, Dutch physicist Heike Kamerlingh Onnes discovered that mercury's resistivity disappears below about 4 Kelvin.

A superconductor is a material that can conduct electricity without resistance when it is cooled below a certain critical temperature, typically near absolute zero. As shown in figure 21.11

This allows charge to flow without energy loss to thermal effects. Superconductors exhibit this behavior below a critical temperature, usually near absolute zero, with resistivity less than $4x10^{-25}\Omega$-m. They also exhibit perfect diamagnetism, expelling magnetic fields in the superconducting state, known as the Meissner effect, and enabling phenomena like magnetic levitation.

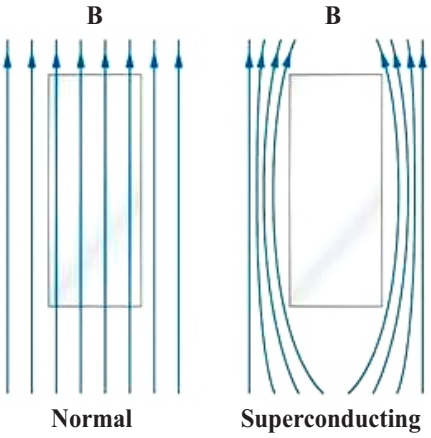


Figure 21.11 supper conductor graph

21.3.4 Super conductor Applications:

Superconductors are used in particle accelerators, generators, transportation, computing, electric motors, medical applications, and power transmission, among other fields.

1. Magnetic Resonance Imaging (MRI):

MRI scanners provide high resolution picture of the tissues inside the body. Superconducting coils produce a strong magnetic field (up to 60,000 times as strong as the intensity of Earth's magnetic field) that is used to align the protons of hydrogen atoms in the body of the

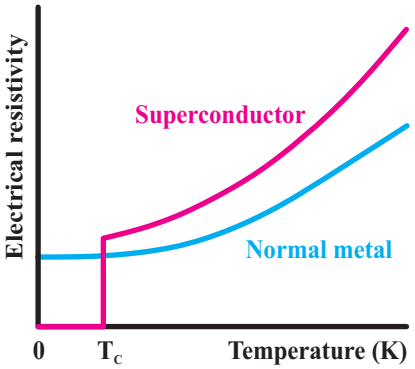


Figure 21.12 Meissner effect super conductors

patient, as shown in figure 21.13. Like electrons, protons have a 'spin' property, so they align with a magnetic field the proton's axis vibrates about the applied magnetic field. Vibrating protons are crashed with a burst of radio waves tuned to push the Protons' spin axes are sideways, perpendicular to the applied magnetic field. When radio waves pass and the protons quickly return to their vibrating pattern, they emit a faint electromagnetic signal whose frequencies depend slightly on the chemical environment in which the proton resides. These signals, detected by sensors, are then analyzed by a computer to reveal varying densities of hydrogen atoms in the body and their interaction s with surrounding tissue. The resulting images clearly distinguish between fluid and bone. MRI was formerly called NMRI (Nuclear Magnetic Resonance Imaging) because hydrogen nuclei resonate with the applied fields.

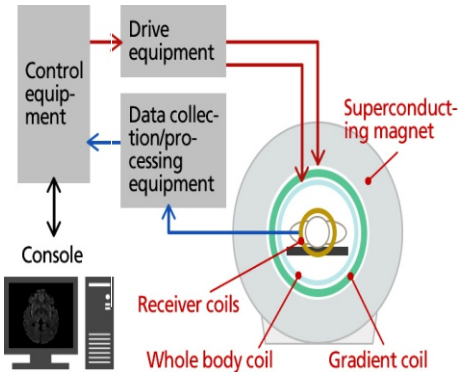


Figure 21.13 working principle of MRI

2. Maglev Train Systems:

Maglev train systems use powerful electromagnets to float the trains over a guide-way, instead of the old steel wheel and track system as shown in figure 21.14. A system called electromagnetic suspension suspends, guides, and propels the trains. A large number of magnets provide controlled tension for lift and propulsion along a track. Maglev (derived from magnetic levitation) is aim of train transportation that uses two sets of magnets: one set to repel and push the train up off the track and another set to move the elevated train ahead, taking advantage of the lack of friction.

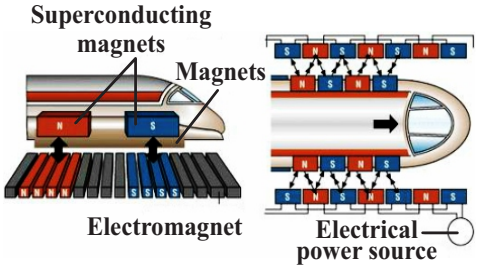


Figure 21.14 Magnetic Levitation

3. Small Electric Motor:

Precision Instruments:

Superconducting motors can be used in precision instruments where high efficiency and low noise are critical.

Medical Devices:

Compact, efficient motors for medical devices such as MRI machines, which already use superconducting magnets.

Aerospace:

High-efficiency, lightweight motors for aerospace applications, where every gram of weight saved is crucial.

Robotics:

Small, powerful motors for advanced robotic applications requiring precise control and high power output.

4. Super Conductivity and Super Computers:

Super conducting components can improve the performance of super computers in several ways as:

Super conducting qubits:

These are the basic building blocks of quantum computers, which are able to perform calculations at speeds that are exponentially faster than classical computers. Superconducting qubits can maintain their quantum states for long periods of time, allowing for complex computations to be performed.

Faster processing speeds:

Super conducting components can process information at much faster speeds than traditional electronic components, resulting in faster computing times and the ability to handle larger datasets.

Lower power consumption:

Super conducting circuits require less power to operate than traditional electronic circuits, reducing energy costs and making them more environmentally friendly.

Improved reliability:

Super conducting components are less prone to errors and can operate for longer periods of time without failure, resulting in more reliable and efficient supercomputers.

However, the use of superconducting components in super computers is still in the experimental stage and faces several challenges.

21.4.1 Magnetic Properties of Solids:

The magnetic properties of solids depend on how their electrons behave in response to an external magnetic field. There are three main types of magnetic materials: ferromagnetic, paramagnetic, and diamagnetic.

Material	Behavior	Examples
Ferromagnetic	Have magnetic moments (tiny magnetic domains) that tend to align parallel to each other. When exposed to an external magnetic field, these materials can become strongly magnetized.	Iron, Cobalt, Nickel.
Paramagnetic	Have individual magnetic moments that are randomly oriented, and they become weakly magnetized when exposed to an external magnetic field	Aluminum Platinum, Titanium
Diamagnetic	Have no intrinsic magnetic moment. When exposed to an external magnetic field, they develop a weak, negative magnetization that opposes the applied field.	Copper, Bismuth.

21.4.2 Magnetic Domain:

The magnetic domain theory was proposed by Weiss in 1907. According to this theory, ferromagnetic material contains a large number of tiny regions, and each region (magnetic domain) consists of atomic magnetic moments that are aligned in a specific direction. Initially, these magnetic domains are randomly oriented, resulting in a weak magnetization where the magnetic moments cancel each other out in the absence of an external magnetic field.

In the presence of a strong external magnetic field, these magnetic domains align in the direction of the field, resulting in strong magnetization. This alignment occurs because the magnetic moments within each domain are parallel due to the influence of the external magnetic field. Paramagnetic, Ferromagnetic and Anti ferromagnetic as shown in Figure 21.15.

(a) Paramagnetic (Randomly oriented order,)
(b) Ferromagnetic (Perfect order)
(c) Anti ferromagnetic (Perfect anti order)

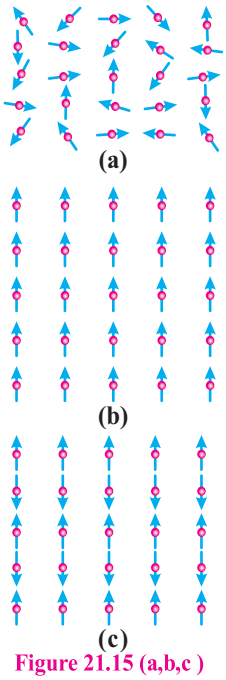


Figure 21.15 (a,b,c) magnetic domain

In ferromagnetic materials, once all the magnetic domains are aligned in one direction and the external field is removed, the domain boundaries do not change their orientation. This property prevents demagnetization, making ferromagnetic materials useful for making permanent magnets. Ferromagnetic materials are strongly attracted to external magnetic fields due to their aligned magnetic domains.

21.4.3 Curie Point:

Curie point is also known as Curie temperature. Some magnetic materials lose their magnetic properties as the temperature rises above the Curie temperature. The Curie temperature weakens the magnetic properties of the material. Pierre Curie, a French physicist, discovered it in 1895 and proposed certain laws related to the magnetic properties of materials as temperature changes.

Curie's Law:

The Curie law states that in a paramagnetic material, the material's ***"magnetization is directly proportional to an applied magnetic field". $M \propto H$***

$M = \chi H$ ..... (H is the applied magnetic field strength)

But when it is heated, the relation is reversed, i.e., the magnetization becomes inversely proportional to temperature.

Mathematically, it is written as:

$$M = \frac{C\,X\,H}{T} \text{.....(21.4)}$$

DO YOU KNOW?

Alloys of iron differ; soft iron is easier to magnetize than steel. It helps to tap the material to nudge any stubborn domains into alignment. Another way is to stroke the material with a magnet. The stroking motion aligns domains. If a permanent magnet is dropped or heated outside the strong magnetic field from which it is made, some of the domains are jostled out of alignment and the magnet becomes weaker.

Where:
M is the magnetization; H is the applied magnetic field, T is the absolute temperature, and C is the Curie constant specific to the material.
The various experiments by Pierre Curie showed that for many substances, the susceptibility χ (chi) is inversely proportional to the absolute temperature T. $\chi = \frac{C}{T}$ .....(21.5)

Consider an example of iron atoms with a temperature of 770°C (1043 K). Each iron atom acts like a tiny magnet at this temperature spontaneously. Each of them will align themselves as some kind of magnetic material.

For pure iron, as shown in Figure 21.16, the atomic magnets are distributed within each microscopic domain. Pure iron is a kind of ferromagnetic material. The directions of the magnetic fields are the same so that their magnetic fields strengthen each other.

Anti-ferromagnetic materials possess atomic magnets that have alternate properties of magnetic fields. They act in opposite directions so that their magnetic fields cancel each other.

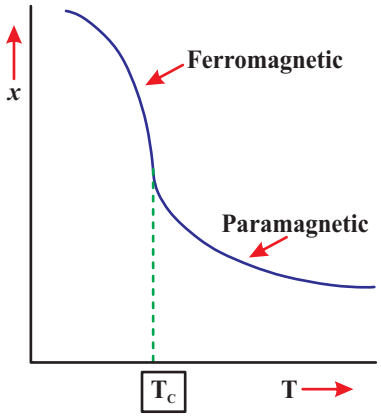


Figure 21.16
critical temperature and curies law

The graph depicts the rise of temperature above the Curie point T_c, which can lead to the production of roughly similar patterns of decreasing magnetic property. This behavior is constant in all three classes of materials.

21.4.4 Hard and Soft Ferromagnetic Substances:

Ferromagnetic materials fall into two main categories: Hard and Soft.

Hard ferromagnetic:

Materials like neodymium, samarium cobalt, and alnico are permanent magnets with high coercivity, suitable for stable magnetic fields in applications like speakers and generators.

Soft ferromagnetic:

Including iron, nickel, and cobalt, these materials have low coercivity, making them easy to magnetize and demagnetize. They are used in devices like transformers and inductors that require rapidly changing magnetic fields. Microstructure influences magnetic properties, with hard substances having a fine-grained structure and soft ones having a coarse-grained structure. The choice between hard and soft ferromagnetic materials depends on the specific application's requirements and operating conditions. For instance, hard ferromagnetic materials are used for permanent magnets in motors, while soft ferromagnetic materials are employed in rotating armatures experiencing rapidly changing magnetic fields.

Table of Curie temperature

TABLE No:21.2
Critical temperature of different elements

Material	Curie T_c(K)
Fe	1043
Co	1388
Ni	627
Gd	293

21.4.5 Hysteresis Loss:

When a material is not magnetized, the axes of different domains point in various directions, resulting in a zero magnetic effect. Upon the application of an external magnetic force, the axes align with the force, creating a strong magnetic field. Repeated cycles of magnetization and demagnetization disturb domain alignment, leading to hysteresis loss. Energy is stored during field establishment and returned during collapse, but hysteresis causes incomplete energy recovery, resulting in heat. This adverse loss, seen in transformers, reduces overall efficiency. To minimize hysteresis loss, soft iron cores are used due to their lower energy loss. Factors affecting hysteresis loss include the area under the hysteresis loop, frequency, and the volume of the material. The loss is directly proportional to these factors, impacting the overall efficiency of devices like transformers.

21.4.6 Synthesis of hysteresis Loop between Magnetic Field Strength and Magnetizing Current.

The hysteresis loop is a graphical representation, as shown in Figure 21.17 that shows the relationship between the magnetic field strength (B) and the magnetization (H) of a material. The loop is particularly useful for understanding how a material responds to changes in an external magnetic field. Here's how you can synthesize the variation of magnetic field strength with magnetizing current from the hysteresis loop.

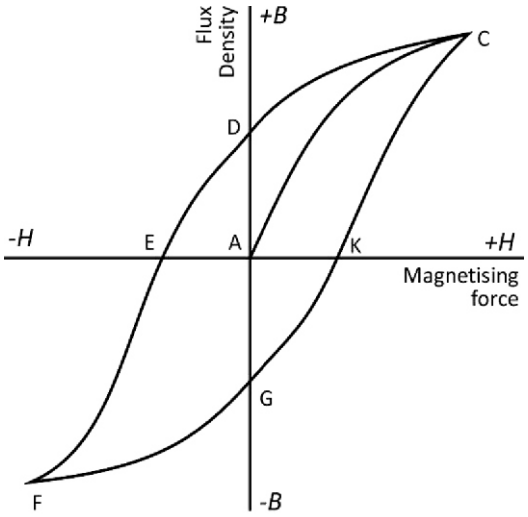


Figure 21.17
hysteresis loop for magnetic material

1. Hysteresis Loop:

A hysteresis loop is a closed curve illustrating a material's magnetic behavior under a changing magnetic field, formed by plotting the material's magnetization against the magnetic field strength.

2. Variation of Magnetic Field Strength with Magnetizing Current:

The horizontal axis of the hysteresis loop represents the magnetizing current or applied magnetic field strength. The vertical axis represents the resulting magnetization of the material.

3. Magnetizing and Demagnetizing:

When the magnetizing current increases, the material gets magnetized, and the magnetic field strength goes up, shown on the right side of the hysteresis loop.

When the magnetizing current decreases, the material may not fully demagnetize, and the magnetic field strength remains at a certain level, shown on the descending left side of the loop.

4. Understanding the Loop:

The width of the hysteresis loop indicates energy loss (hysteresis loss) during magnetization and demagnetization cycles.

The shape of the loop provides information about the material's magnetic properties, such as coercivity and remanence.

By analyzing the hysteresis loop, the relationship between magnetic field strength and magnetizing current can be observed, aiding in understanding the material's magnetic properties and response to external magnetic field changes.

21.4.7 Hysteresis loop for permanent and temporary magnets:

Hysteresis is especially marked in materials of high residual magnetism, such as hardened steel. In most cases, hysteresis is a necessity as it causes dissipation of heat, waste of energy, and vibration due to change in polarity and rotation of element magnets in the material.

Hysteresis loops for hard steel, wrought iron, cast steel, and for alloyed sheet steel are shown in Figure 21.18.

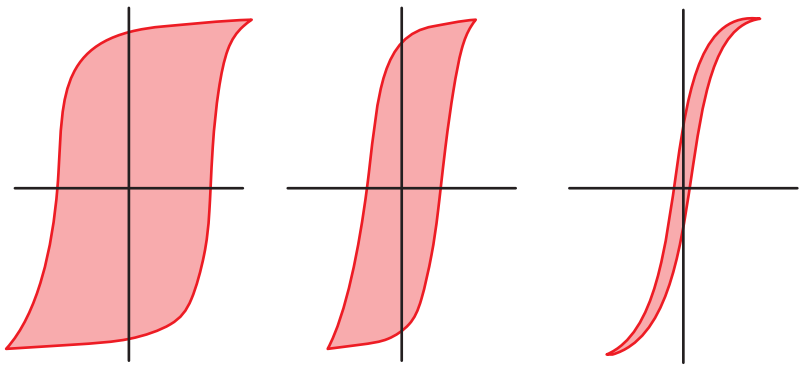


Figure 21.18(a, b, c)
hysteresis loops for hard steel, wrought iron and cast steel and low carbon steel respectively

Loop(a) is for hard steel. Due to its high retentivity power and large coercive force, this material is well-suited for permanent magnets. Since the area of the hysteresis loop for hard steel is large, hard steel is not suitable for rapid reversals of magnetization. As shown in Figure 21.18(a)

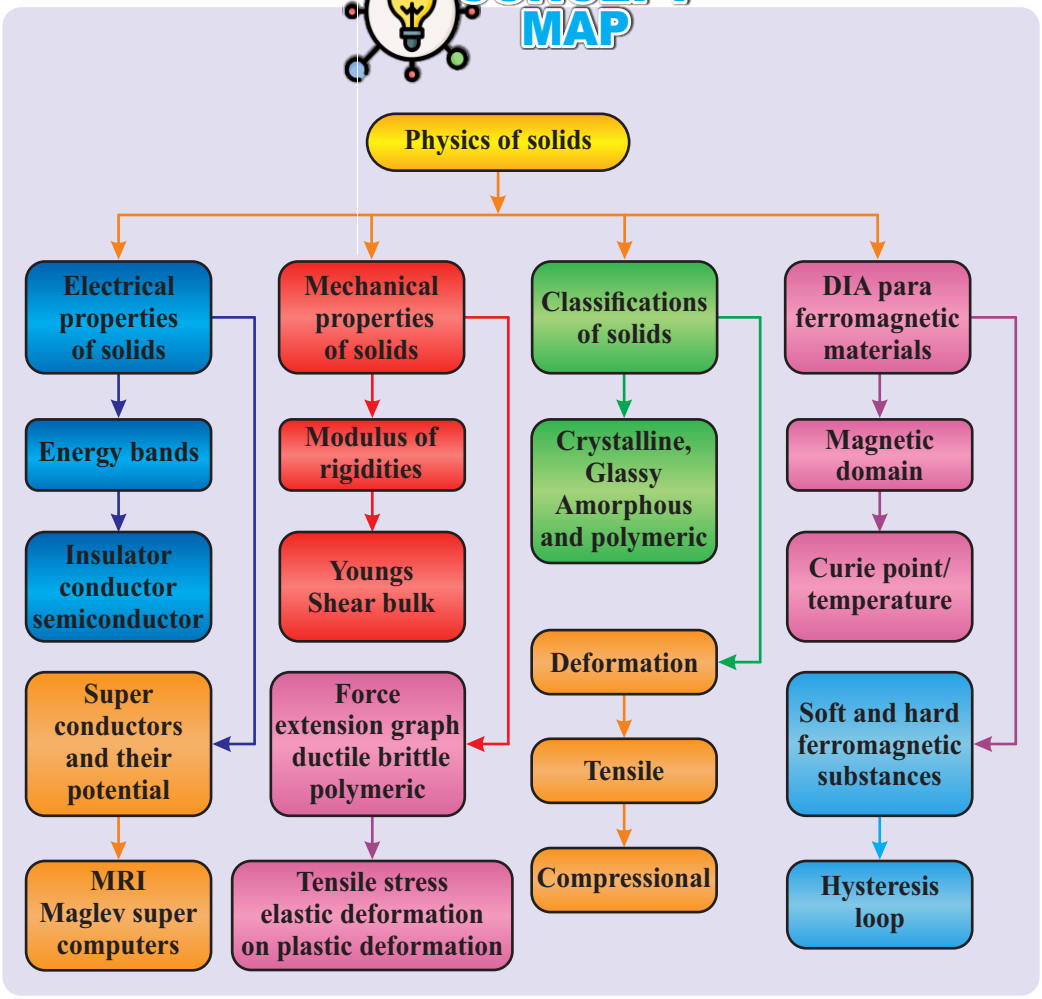
Loop (b) is for wrought iron and cast steel, which rise steeply. Hence, these materials have high magnetic permeability and good retentivity; therefore, these materials are suitable for cores of electromagnets. As illustrated in Figure 21.18(b)

Loop(c) is for iron, low carbon steel, silicon alloys, perm-alloy (alloy of Nickel and Iron), or Mu-metal (Nickel-Iron soft ferromagnetic material) sheets. Since the permeability of these materials is very high and hysteresis losses are very low, these materials are most suitable for transformer cores and armatures, which are subjected to rapid reversals of magnetization. As shown in Figure 21.18(c).



- ✓ A crystalline solid is one where the internal structure is regular in nature, the atoms within it being set in well-defined patterns.
- ✓ Amorphous materials do not possess repeated structures over a long range as in crystalline solids. Rubber is an example of a polymer. It consists of very long chain molecules.
- ✓ Stress is a quantity which defines the magnitude of forces that cause deformation. Stress is usually defined as force per unit area.
- ✓ When forces pull on an object and cause its elongation, like the stretching of an elastic band, we call such stress a tensile stress.
- ✓ Strain is given as a fractional change in either length (under tensile stress) or volume (under bulk stress) or geometry (under shear stress).
- ✓ The tensile strength of a solid is a measure of its resistance to being pulled apart.
- ✓ Ductility is a property of metal that enables it to be drawn through a die to produce a wire.
- ✓ Elasticity is a measure of a deformed object's ability to return to its original size and shape once the outside forces are removed.
- ✓ Young's modulus is defined as the ratio of stress to strain.
- ✓ Bulk Modulus If an object is subjected to inward forces from all sides, its volume will decrease.
- ✓ Energy Band: The electrons in the same orbit exhibit different energy levels. The grouping of these different energy levels is known as an energy band.
- ✓ Superconductivity allows charge to flow through a superconducting conductor without losing its energy to thermal energy.
- ✓ Magnetic Hysteresis / Retentivity: When a magnetic material is placed in an external magnetic field due to which its grains get oriented in its direction and the material is magnetized. If this material is removed from that external magnetic field, even some magnetization remains in that material which is called residual magnetism. This is also called retentivity of the material.
- ✓ A magnetic domain is a region within a magnetic material in which magnetization is in a uniform direction.
- ✓ Curie point or Temperature is defined as when some magnetic materials lose their magnetic property as temperature becomes above the Curie temperature.
- ✓ Hysteresis Loss is the disturbance in the alignment of the various domains causes hysteresis loss.

CONCEPT MAP



EXERCISE

Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. The property of a body by virtue of which it tends to regain its original size and shape when the applied force is removed is called
   (a) Elasticity (b) Plasticity (c) Rigidity (d) Compressibility
2. Substances which can be stretched to cause large strains are called
   (a) Brittle (b) Ductile (c) Plastic (d) Elastomer

3. If the load is increased beyond the point, the strain increases rapidly for even a small change in the stress.
   (a) Elastic point (b) Yield point (c) Plastic point (d) Fracture point
4. The reciprocal of the bulk modulus is called
   (a) Compressibility (b) Volume stress
   (c) Modulus of rigidity (d) Volume strain
5. Which of the following statements is/are wrong?
   i. Hollow shaft is much stronger than a solid of the same length and same mass.
   ii. Reciprocal of bulk modulus of elasticity is called compressibility.
   iii. It is difficult to twist along rod as compared to small rod.
   (a) III only (b) I only (c) II and III (d) I and II
6. Metals are good conductors of heat and electricity. This property is conferred by
   (a) Covalent (b) Ionic (c) Metallic (d) Hydrogen
7. For a metallic crystal, delocalized electrons occupied band is:
   (a) Conduction band (b) Valence band
   (c) Conduction and valence bands (d) there are no delocalized electrons
8. The semiconductors have resistivity
   (a) Between conductor and insulator (b) More than insulators
   (c) Less than conductors
   (d) Depending upon the semiconductor material property.
9. A ferromagnetic substance becomes a permanent magnet when it is placed in a magnetic field because
   (a) All the domains get oriented in the direction of the magnetic field.
   (b) All the domains get oriented in the direction opposite to the direction of the magnetic field.
   (c) Domains get oriented randomly.
   (d) Domains are not affected by the magnetic field.
10. Identify methods to demagnetize a ferromagnet.
    (a) By cooling, heating, or submerging in water
    (b) By heating, hammering, and spinning it in an external magnetic field
    (c) By hammering, heating, and rubbing with cloth
    (d) By cooling, submerging in water, or rubbing with cloth

Section (B): CRQs (Short Answered Questions):

1. Why are the springs made of steel and not of copper?
2. The breaking force for a wire is F. What will be the breaking force for two parallel wires of the same size?
3. Distinguish between intrinsic and extrinsic semiconductor.
4. A wire is replaced by another wire of same length and material but of twice the diameter. What will be the effect on the:
   (a) Increase in its length under a given load? (b) Maximum load which it can bear?
5. Sand does not possess any definite shape and volume, still it is solid. Give reason.
6. Specify the importance of stress-strain curve.
7. Why liquids don't possess rigidity?

8. Give applications of Curie point.
9. What are amorphous materials and what are their uses?

Section (C): ERQs (Long Answered Questions):

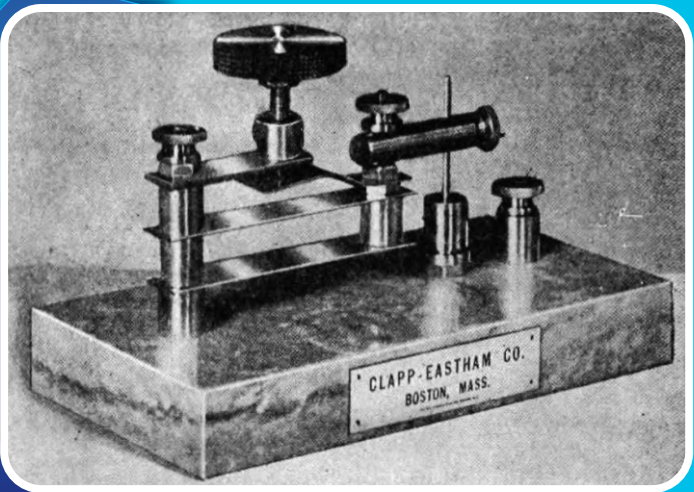
1. Explain Force-Extension graph.
2. Derive relation for Young's Modulus and Shear Modulus.
3. Distinguish between structure of crystalline, glassy, amorphous, and polymeric solids.
4. Describe the energy bands in solids.
5. Describe superconductivity and its applications.
6. Discuss the applications of superconductors for MRI, Maglev's, and supercomputers.
7. Describe hysteresis loss.
8. Synthesize hysteresis loop for relationship between magnetic field strength and magnetizing current.
9. Discuss energy bands and their classification. Explain magnetic properties of soft and hard magnetic materials.

Section (D): Numerical:

1. The 'lead' in pencils is a graphite composition with a Young's modulus of 1.0×10^9 N/m^2. Calculate the change in length of the lead in an automatic pencil if you tap it straight into the pencil with a force of 4.0 N. The lead is 0.50 mm in diameter and 60 mm long. **(1.0mm)**
2. A wire of 2.2 m long and 2.25 mm in diameter, when stretched by a weight of 8.8 kg, its length has been increased by 0.25 mm. Find the stress, strain, and Young's modulus of the material of the wire. Given g = 9.8 m/s^2. **(2.2×10^7 N/m^2, 1.14×10^{-4}, 2×10^{11} N/m^2)**
3. A farmer making juice fills a glass bottle to the brim and caps it tightly. The juice expands more than the glass when it warms up, in such a way that the volume increases by 0.2% (i.e., $\frac{\Delta V}{V_0} = 2 \times 10^{-3}$) relative to the space available. Calculate the normal force exerted by the juice per square centimeter, if its bulk modulus is 1.8×10^9 N/m^2. Assuming that the bottle does not break. **(432 N/cm^2)**
4. The elastic limit of copper is 1.5×10^8 N/m^2. It is to be stretched by a load of 10 kg. Find the diameter of the wire if the elastic limit is not to be exceeded. **(0.912 mm)**
5. What would be the greatest length of a steel wire which is fixed at one end, and can it be hanged freely without breaking? The breaking stress of steel is 7.8×10^8 N/m^2, and the density of steel is 7800 kg/m^3. **(1.02×10^4 m)**
6. A mild steel wire of radius 0.55 mm and length 3.5 m is stretched by a force of 52 N. Calculate: (a) Longitudinal stress, (b) Longitudinal strain, and (c) Elongation produced in the wire if Young's modulus is 2.1×10^{11} N/m^2. **($5.47 \times 10\text{^}7$ N/m^2, 2.6×10^{-4}, 0.91 mm)**
7. Calculate the change in volume of a lead block of volume 1.3 m^3 subjected to a pressure of 12 atm. Also, calculate the compressibility of lead. Given the bulk modulus as B=80×10^9 N/m^2. **(1.97×10^{-5} m^3, 1.25×10^{-11} N/m^2)**
8. The thickness of a metal plate is 0.35 inches. It's drilled to have a hole of radius 0.08 inches on the plate. If the shear strength is 4×10^4 lbs/in^2, determine the force needed to make that hole. **(0.176 in^2, 7×10^3lbs)**

Unit 22 Solid State Electronics

Teaching Periods 15 **Weightage %** 10



The cat whisker device, an early solid-state electronic component, was discovered by German physicist Ferdinand Braun in 1874. Braun found that by touching a thin wire (the "whisker") to a crystal of galena (lead sulfide), he could detect radio waves, creating a primitive radio detector. This discovery led to the development of crystal detectors, used in early radios, and paved the way for modern semiconductor devices, such as diodes and transistors, which revolutionized electronics.

In this unit student should be able to:

- Intrinsic (pure) and doped semiconductors.
- How the *N*-type and *P*-type semiconductors are produced.
- Explain the concept of holes and electrons in semiconductors.
- Explain how electrons and holes flow across a junction.
- Describe a *P-N* junction(diode).
- Discuss its forward and reverse biasing.
- Describe the I-V characteristic curves of *P-N* junction.
- Define rectification and describe the use of diodes for half and full wave rectifications.
- Describe the function and use of LED, Photodiode and Photo voltaic cell.
- Distinguish between *P-N-P*&*N-P-N* transistors.
- Describe the operations of transistors.
- Deduce current equation and apply it to solve problems on transistors.
- Apply operation principles of the transistor including I-V characteristics and biasing methods.
- Explain the use of transistors as a switch and an amplifier (common-emitter).
- Explain common-base and common collector configurations.
- Describe the properties of an ideal operational amplifier.
- Express operational amplifier as a comparator.
- Understand the effects of negative feedback on the gain of an operational amplifier.
- Draw the circuit diagrams for both the inverting and the non-inverting amplifier for single signal input.
- Understand the virtual earth approximation and derive an
- Express for the gain of inverting amplifiers.
- Recall and use expressions for the voltage gain of inverting and of non-inverting amplifiers.

Introduction:

Solid-state electronics is a branch of electronics that deals with the flow of electric current through solid materials, such as semiconductors, insulators, and conductors. It involves the design, development, and application of electronic devices and circuits that use these materials to control and manipulate electrical energy. Solid-state electronics has revolutionized the field of electronics, enabling the creation of smaller, faster, and more efficient devices, such as transistors, diodes, integrated circuits, and microprocessors, which are essential components in modern computers, smart phones, and other electronic devices. The field is constantly evolving, with advancements in materials science, nanotechnology, and quantum mechanics.

22.1.1 Intrinsic (Pure) and Extrinsic (Doped) Semiconductors:

Intrinsic (Pure) Semiconductors:

Intrinsic semiconductors are materials that are chemically pure, meaning they do not contain any significant amount of impurity atoms. The two most commonly used intrinsic semiconductors are silicon (Si) and germanium (Ge). These materials have a crystalline structure, and their electrical properties are determined by the arrangement and behavior of their atoms.

Extrinsic (doped) Semiconductors:

Doped semiconductors are created by introducing impurity atoms into the intrinsic semiconductor material. This process is known as doping, and it significantly alters the electrical properties of the semiconductor. Doping can be done with two types of impurities, resulting in two types of doped semiconductors: n-type and p-type.

In summary, intrinsic semiconductors are pure materials with moderate conductivity, whereas doped semiconductors are intentionally infused with impurities to enhance their electrical properties, making them essential for modern electronic devices.

22.1.2 Production of P-Type and N -Type Semiconductors:

The intentional addition of impurities alters the conductivity and electrical properties of the semiconductor, allowing for the controlled operation of charge carriers.

P-type Semiconductors:

It is characterized by an excess of positively charged carriers, known as "holes," or majority Positive charge carriers.

In *P*-type semiconductors, the holes effectively behave as positive charge carriers in the material. These holes are essentially vacancies in the crystal lattice structure, where an electron is missing from a **covalent bond**.

Semiconductor material, silicon (Si) or germanium (Ge) is the base semiconductor material.

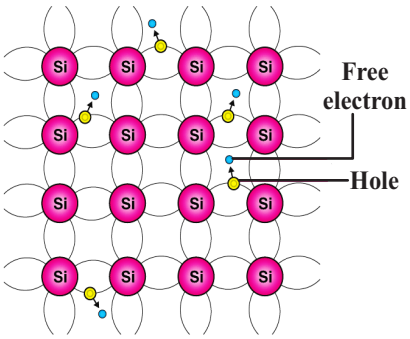


Figure 22.1
free electrons and hole in semiconductor

Doping with Donor Impurities:

A small amount of a group-V element (penta-valent) is doped into the semiconductor's crystal lattice, as illustrated in figure 22.2. Common donor impurities include phosphorus (P) or arsenic (As).

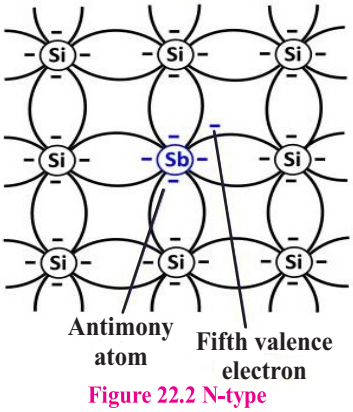


Figure 22.2 N-type

DO YOU KNOW?

Acceptor Ionization: At normal operating temperatures, thermal energy is sufficient to ionize the acceptor atoms, creating mobile holes in the valence band. These holes contribute to the overall conductivity of the material.

These impurity atoms have one more valence electron than the semiconductor material, creating extra electrons that are free to move and conduct electricity.

N-type Semiconductors:

It is characterized by an excess of negatively charged carriers, namely electrons, or majority negative charge carriers.

In *N*-type semiconductors, electrons are the primary charge carriers responsible for the material's electrical conductivity.

Doping with Acceptor Impurities:

If element of group-III (tri-valent) is added into the crystal lattice of the semiconductor as shown in figure 22.3. Common acceptor impurities include boron (B) or gallium (Ga). These impurity atoms have one fewer valence electron than the semiconductor material, creating "holes" in the crystal lattice where electrons can move.

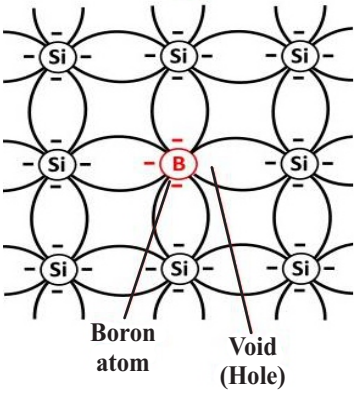


Figure 22.3 P-type

22.1.3 Electron and Hole in Semiconductor:

When an electric field is applied then electric current flow is performed in two ways.

When the temperature is increased then some ***valence electrons*** **enter into the conduction band** and become free electrons. Whenever an electron enters into the conduction band, then a vacant space or a hole is produced in the valence band as shown in figure 22.4. Due to the influence

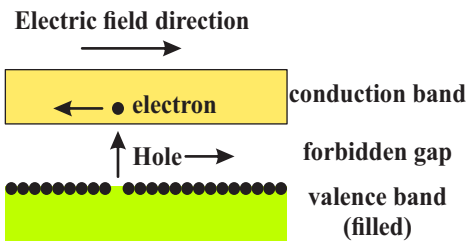


Figure 22.4 electrons and Holes

of potential difference or electric field, flow of electrons and holes are produced.

At room temperature, some electrons from the valence band can cross this small energy gap and are accumulated in the conduction band. As a result, an equal number of holes and electrons are produced. This phenomenon is called electron-hole pair Production.

At the temperature above absolute zero, if the electric field is applied then conduction electrons move towards anode and holes move towards the cathode. So, it can be said that the current flow in the semiconductor is the flow of electrons and holes in the valence and conduction bands in opposite directions.

DO YOU KNOW?

Most familiar intrinsic semiconductors are germanium and silicon. The energy gap between their valence band and conduction band are 0.72 eV and 1.1 eV respectively.

22.2 P-N Junction:

The ***P-N*** junction is formed when the *p*-type and *n*-type semiconductors are joined, is called as *P-N* junction as shown in figure 22.5. *P-N* junction is formed within a single crystal of material rather than just simply joining or fusing together two separate pieces. This junction is a crucial component in electronics and forms the basis for many semiconductor devices like diodes and transistors.

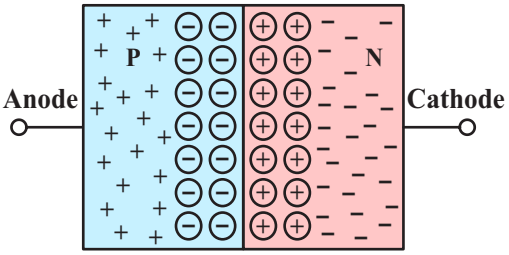


Figure 22.5PN JUNCTION

22.2.1 Flow of Electrons and Holes across Junction:

Initially, a large density rise exists across the junction who leads to a migration of electrons and holes. When electrons move from the N-type to the P-type, negative ions are left behind. Simultaneously, holes from P-type migrate toward the N-type, creating positive ions. This process which is called diffusion, continues until an equilibrium is reached, resulting in a potential barrier zone around the junction, as shown in figure 22.6.

The potential barrier, formed due to repulsion between donor and acceptor atoms, creates a depletion layer. This layer is devoid of free charge carriers, preventing further movement across the junction.

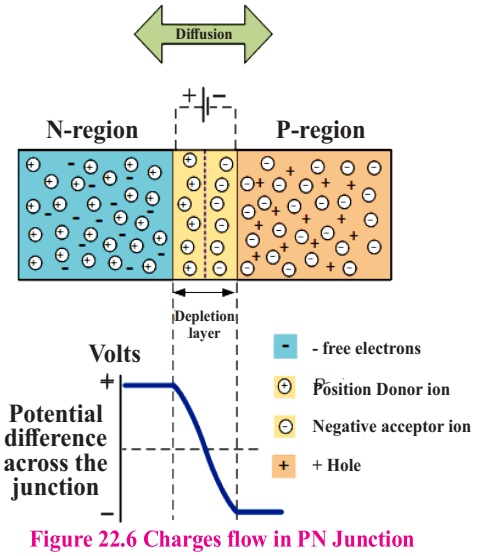


Figure 22.6 Charges flow in PN Junction

22.2.2 P-N Junction as Diode:

A diode is a semiconductor device that is formed through P-N junction and used in allowing the flow of electric current in one direction and blocking in the opposite. The symbol of the *P-N* junction diode is shown in figure 22.7

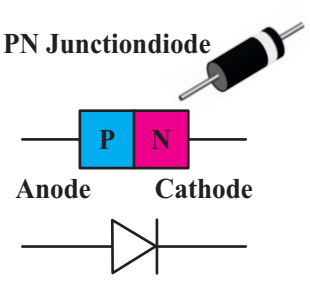


Figure 22.7 Diode

Properties of Diode:

Below are some of the common properties of a diode:

1. Diode has the ability to rectify electric current.
2. It can create a potential barrier and make use of its capacitance properties.
3. Diode creates various nonlinear current-voltage characteristics.

22.2.3 Diode Biasing:

Forward Bias:

When the P-type is connected to the positive terminal of the battery and the N-type is connected to the negative terminal is called Forward bias as shown in figure 22.8.

In this condition, the applied electric field and the built-in electric field at the *P-N* junction are in opposing directions. Adding both the electric fields gives a resultant electric field, which is found to be smaller than the built-in electric field. So the depletion region becomes thinner and less resistant.

When the applied voltage is high, the resistance of the depletion region becomes insignificant. At 0.3V to 0.6 V, the resistance of the depletion region in silicon becomes absolutely insignificant, allowing current to flow freely through it.

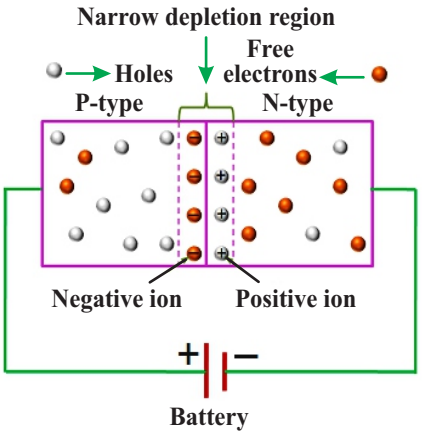


Figure 22.8 Forward Bias

Reverse Bias:

When the P-type is connected to the negative terminal of the battery and the N-type is connected to the positive side is called Reverse Bias as shown in figure 22.9

In this condition, the applied electric field and the built-in electric field are both in the same direction. The resultant electric field and the built-in electric field are also in the same direction, resulting in a more resistive, thicker depletion region. If applied voltage increased, it results thicker and more resistant depletion region.

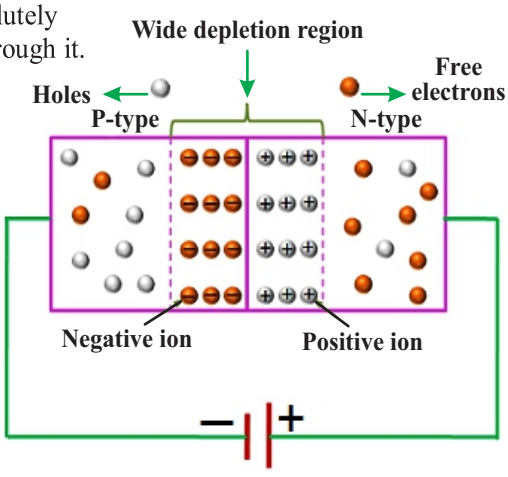


Figure 22.9 Reverse Bias

Self-Assessment Questions:

1. What causes the formation of the depletion region in a PN junction at equilibrium?
2. How does forward biasing affect the depletion region and current flow in a PN junction?
3. Why is the current very small in a PN junction under reverse bias?

22.2.4 I-V Characteristics of *p-n* Junction:

The relationship between the voltage across the junction and current through the circuit is known as the (V-I) characteristics of a *P-N* junction or semiconductor diode.

The V-I characteristics of the *P-N* junction can be explained in three cases:

- Zero bias or unbiased
- Forward bias
- Reverse bias

At zero bias, no electric current flows through the diode because there's no external voltage applied to enable the movement of electrons or holes. In forward bias as shown in figure 22.10 (a) when the diode voltage (V_d) reaches 0.7 V for silicon and 0.3 V for germanium, current starts flowing. The current increases gradually at first, creating a non-linear curve until the diode surpasses the potential barrier, after which it operates normally and the curve steepens linearly with increasing external voltage.

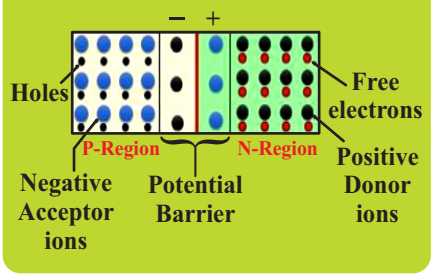
In reverse bias, only a small leakage current flows, represented to the left of the origin in the graph as shown in figure 22.10(c). This current remains low until the diode breaks down, at which point it can be destroyed unless a high series resistance limits the current.

DO YOU KNOW?

The depletion layer or region acts as a barrier that controls the flow of current. When a voltage is applied, the depletion layer widens or narrows, allowing more or fewer charge carriers to pass through. This mechanism enables the transistor to amplify or switch electronic signals.

The depletion layer is characterized by:

1. Reduced carrier concentration
2. Increased resistance
3. Decreased conductivity



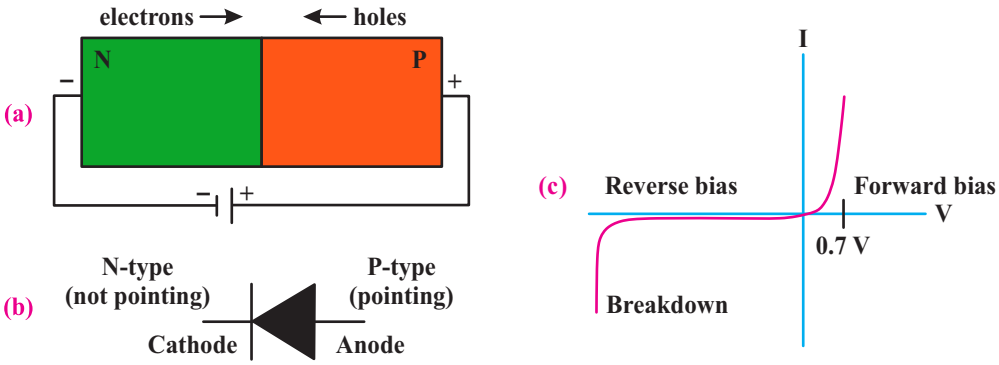


Figure 22.10 I-V Characteristics of *p-n* Junction (a,b,c)

Self-Assessment Questions:

1. Why does reverse saturation current occur in a PN junction?
2. How do the IV characteristics of a PN junction change in forward bias?

22.2.5 Rectification:

Rectification is the process of converting an alternating current (AC) waveform into a direct current (DC) waveform, i.e., creating a new waveform that has only a single polarity as shown in 22.11.



Figure 22.11 Rectification

A diode allows electric current to flow in forward bias condition and blocks the current in reverse bias condition.

Rectification is classified into two types according to the output characteristics which are:

(i) half-wave rectification and
(ii) full-wave rectification.

Half Wave Rectification:

Since a diode allows ***AC*** current to flow only in one direction, it can serve as a rectifier. As shown in figure 22.12(a). The *AC* source applies a voltage across the diode alternately positive and negative. When the positive cycle of AC voltage passes through the diode, the diode is forward biased and act as a closed circuit there is current through the resistor R.

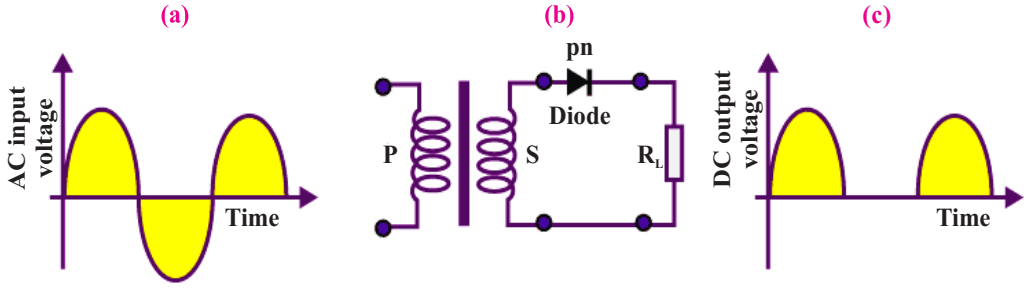


Figure 22.12 (a,b,c) Half Wave Rectifications

During the negative half cycle, the diode is connected with the negative supply which reverses biases the diode, the diode behaves like an open circuit and does not produce the output across the load.

Hence a graph of the voltage V_{ab} across R as a function of time looks likes the output voltage shown in figure 22.12(b). This is called half wave rectification which seems unidirectional from its output signal.

Applications of a Half-wave Rectifier:

- Low power simple battery charger circuit.
- Fire Alarm circuits.
- Soldering Iron circuit.
- Amplitude Modulation (AM) Radio circuits as a Detector.

Full-Wave Rectification:

Full-wave rectification uses two diodes (as shown in figure 22.13(a,b,c).During the positive half cycle of input AC signal, this makes the diode D_1 forward biased (acts as closed switch) and diode D_2 reverse biased (acts as open switch). Therefore, current flows through the load resistor R.

During the negative half cycle of input AC signal, this makes the diode D_2 forward biased and the diode D_1 reverse biased. Therefore, the current will flow through diode D_2 through, load resistor R and lower half of the secondary winding.

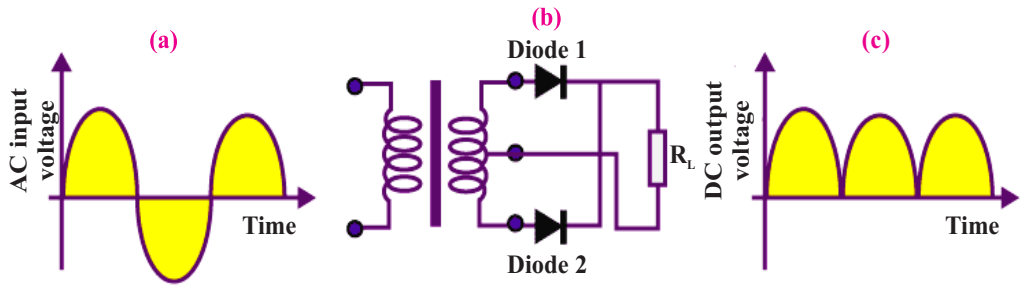


Figure 22.13 (a,b,c) Full-Wave Rectifications

Note that the current through the load is in the same direction for both half cycles of input AC supply. Hence DC output is obtained across the load.

Applications of a Full-wave Rectification:

- Mobile phones, laptops, charger circuits.
- Uninterruptible Power Supply (UPS) circuits to convert AC to DC.
- Our home inverters convert AC to DC.
- LCD, LED TVs.

22.2.6 Function and uses of light emitting diode (LED), Photodiode, Photo voltaic cell:

1. Light Emitting Diode (LED):

A light-emitting diode (LED) is a special type of junction in which current flows, when it is activated in a forward direction as shown in figure 22.14.

In this process, electrons move from the negative side to the positive side, combining with holes emits a photon with energy similar to the material's band gap.

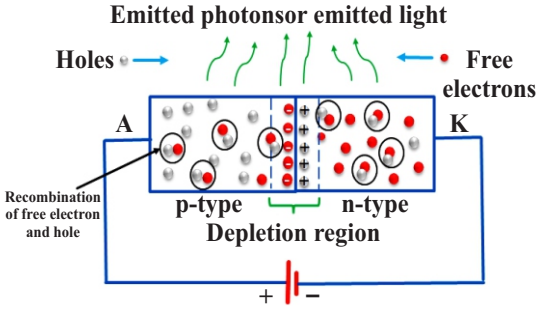


Figure 22.14 Light emitting diode

Silicon diodes don't work well for this, but compounds like gallium and arsenic create efficient LEDs. Gallium arsenide (GaAs), with a crystal structure similar to silicon, is often used..

LED bulbs replace traditional lighting in various applications such as flashlights, streetlights, traffic signals, car brake lights, billboards, and LCD screens. LEDs, also known as solid-state lighting, last longer (50000 hours vs 2000 hours for regular bulbs), are more efficient, and durable.

DO YOU KNOW?

Doping GaAs with group VI atoms like Se and group II atoms as acceptors enhances its properties. GaAs has a 1.42 eV energy gap, producing near-infrared photons (870 nm wavelength), making it suitable for infrared LEDs in remote controls for devices like TVs, DVD players, and car door locks

2. Photodiode:

A photodiode is a type of light detector that converts light into current or voltage. It includes optical filters, built-in lenses, and surface areas.

Photodiodes are often used in reverse bias, where a voltage encourages the flow of photocurrent, maximizing sensitivity. These detectors are usually made of semiconductor materials like silicon, which can absorb photons of light.

When light particles (photons) hit the semiconductor material, they transfer energy to electrons, creating electron-hole pairs. This process generates an electric current proportional to the incident light's intensity.

In simple terms, brighter light results in a higher current from the photodiode.

Photodiodes find widespread use in various applications, including:

- **Optical Communication:** Photodiodes are employed in optical communication systems, such as fiber optics, to detect and convert transmitted light signals into electrical signals.
- **Light Sensors:** They are used in electronic devices like cameras, light meters, and automatic lighting systems to sense ambient light levels.
- **Barcode Readers:** Photodiodes are often used in barcode scanners to detect the reflected light from the barcode.
- **Medical Devices:** In some medical instruments, photodiodes are used for tasks like measuring oxygen levels in blood.

3. Photo Voltaic Cells:

Solar cell or photovoltaic cell is heavily doped P-N junction diodes used to convert sunlight into electric energy.

If the energy of incident photon is greater than the band gap energy, it creates electron-hole pairs, as shown in figure 22.15. That absorbed photon excites an electron from the valence band and produce a current, when connected to an external circuit, work as source of emf and power. The basic element of a photovoltaic cell is a

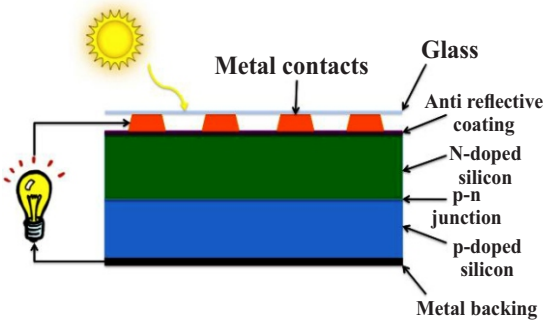


Figure 22.15 Photo Voltaic Cells

semiconductor material, usually made of silicon. Silicon is chosen for its widespread availability. Other semiconductor materials, such as cadmium telluride or copper indium gallium selenide, are also used in different types of solar cells.

A typical silicon *P-N* junction may produce about 0.6 V. Many are connected in series to produce higher voltage. Such series strings are connected in parallel within a photovoltaic panel i.e., solar panel. A good photovoltaic panel can have an output of perhaps 50 W/m^2, averaged over day and night, sunny and cloudy.

Uses of Photodiodes:

- **Camera:** Light meters, Automatic Shutter Control, Photographic Flash Control
- **Medical:** CAT Scanners, X-Ray Detection, Pulse Oximeters, Blood Particle Analysis
- **Automotive:** Headlight Dimmer, Twilight Detectors
- **Communication:** Fiber Optic Link, Optical Remote Control
- **Industry:** Bar Code Scanners, Light Pen, Encoders, Surveying Instruments, Copiers-Density of Toner.

22.3 Transistors and their characteristics:

A transistor is a semiconductor device that amplifies or switches electronic signals by controlling the flow of current between its three layers: base, collector, and emitter. It acts as a gatekeeper, allowing or blocking the flow of current between the collector and emitter, based on the voltage applied to the base.

Both *N-P-N* and *P-N-P* transistors can be made, and they are shown schematically in figure 22.16.

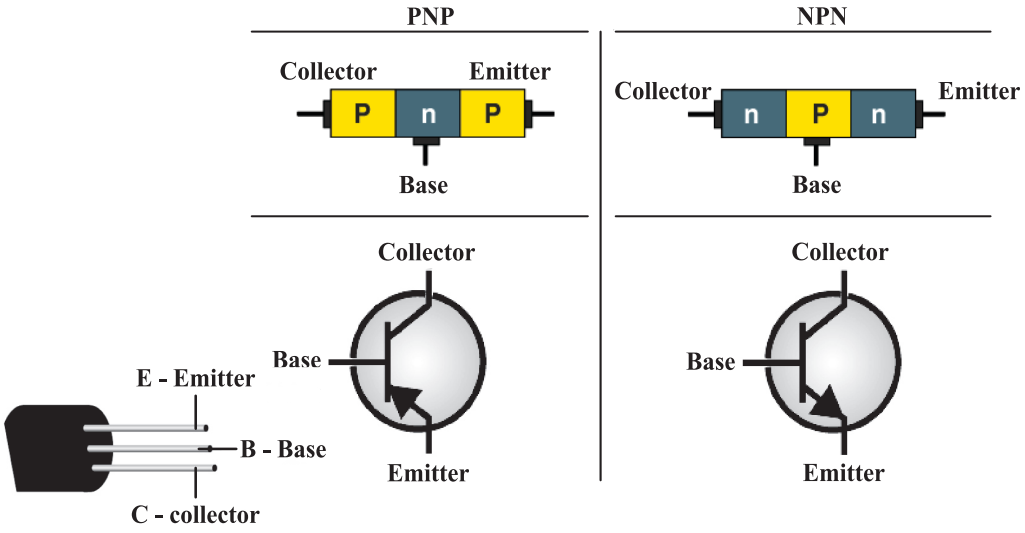


Figure 22.16 (a) Transistor

Figure 22.16 (b) PNP and (c) NPN JUNCTION

The arrow is always placed on the emitter; it indicates the direction of current flow in normal operation.

There are three main characteristics listed below:

1. Input Characteristics
2. Output Characteristics
3. Constant-Current Characteristics.

22.3.1 Distinguish between *N-P-N* and *P-N-P*:

N-P-N	*P-N-P*
It contains two N-type semiconductors.	It contains two P-type semiconductors.
Its current is directed outwardly.	Its current is directed inwardly.
In this transistor combination, electrons are majority current carriers.	In this transistor combination, holes are majority current carriers.
Its current direction is emitter to base.	Its current direction is base to emitter.
It has holes as minority current carriers.	It has electrons as minority carriers.
It's switched on when electrons enter the base.	It's switched on when holes enter the base.
As current reduces at base, transistor does not function across the collector terminal and switches off.	When current is available at base, then transistor switches off.

22.3.2 Operation of transistors:

Transistors are categorized as PNP and NPN. For simplicity the NPN transistor is taken as shown in figure 22.17 (a). Its action has two requirements:

1. The base-emitter junction is forward biased so that a current I_B is generated. Once this junction is conducting, $V_{BE} \cong 0.6$ V
2. The base-collector junction is reverse biased.

Transistor action then translates into relationship

$$I_C = \beta I_B \ldots\ldots\ldots\ldots\ldots 22.1$$

Where β is the current gain, which is typically 100. For completeness, the emitter current is then given by

$$I_E = I_B + I_C = (1+\beta)I_B \ldots\ldots\ldots\ldots\ldots 22.2$$

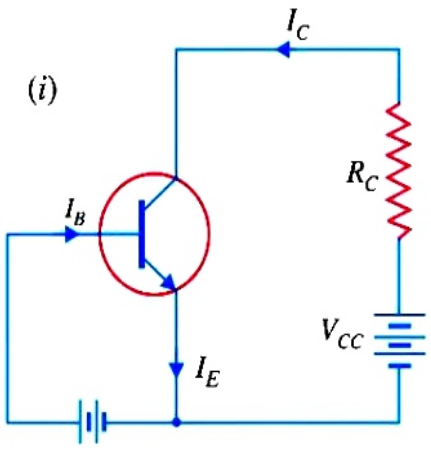


Figure 22.17 (a) NPN transistor

Figure 22.17 graphs a set of characteristics curves of the transistor. Those graphs show the collector current Ic as function of the collector-emitter voltage V_{CE}, for different values of the base current I_B. When the transistor-action conditions are not met, the transistor is off and Ic =0. This situation is called ***cutoff.***

Now assume that V_{BE} is forward biased and I_B is above some minimum value. As the collector-base junction becomes reverse biased, which we represent here by increasing V_{CE}, transistor action begins to unfold:

In a transistor's char acteristic curve, I_c (collector current) initially rises sharply until it reaches a certain point. Beyond this point, I_c continues to increase, but at a much slower rate. This might seem counterintuitive, but the rapid rise is called "saturation," indicating full current flow. The flat region that follows is when the transistor operates normally, where further increases in base current do not significantly increase I_c

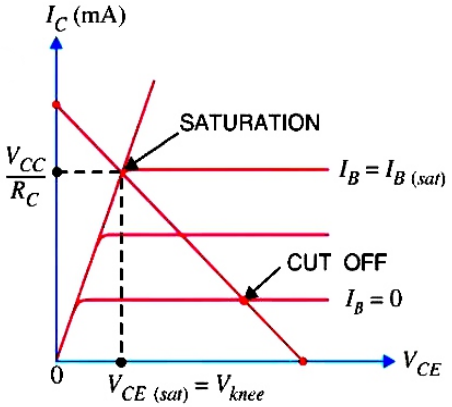


Figure 22.17 (b) transistor characteristics curves

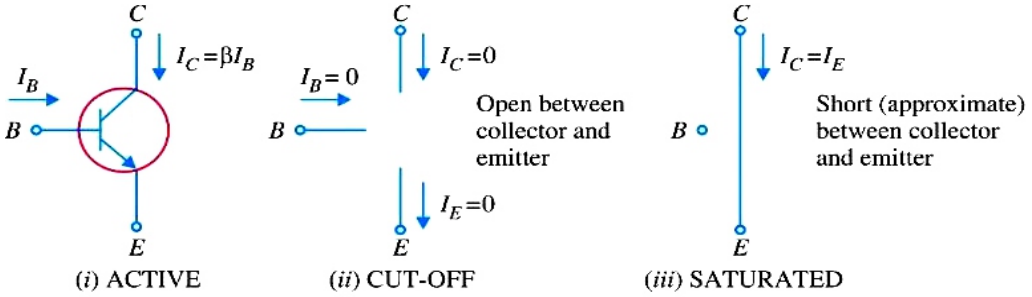


Figure 22.18 Transistors active-cut off and saturated

22.3.3 Transistors current equation:

It is a three-terminal device, with three possible ways to connect it within an electronic circuit with one terminal being common to both the input and output signals.

Bipolar Junction Transistor Configurations:

Each method of connection responding differently to its input signal within a circuit as the static characteristics of the transistor varies with each circuit arrangement.

The Common Emitter Amplifier Circuit

In this circuit shown in Figure 22.19, the current leaving the transistor must be equal the currents entering it because the emitter current (I_E) is the sum of the collector current (I_C) and the base current (I_B).

The common emitter transistor configuration has a substantial current gain because

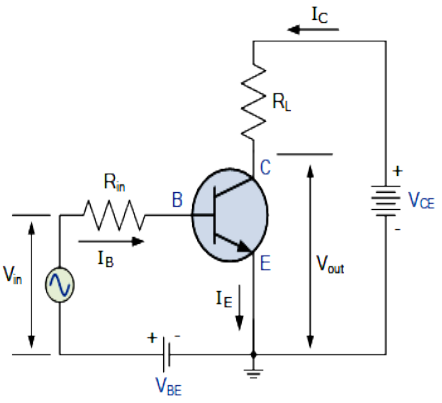


Figure 22.19 Common Emitter Amplifier

the load resistance (R_L) is connected in series with the collector. The current gain is represented by Beta (β),which is the ratio of I_C to I_B. For the common emitter setup,

$$I_E = I_C + I_B, \ldots\ldots\ldots\ldots 22.3$$

where the ratio of I_C to I_E is known as Alpha (α).It's important to note that Alpha is always less than one.

Due to the physical construction of the transistor, any small change in the base current (I_B) leads to a much larger change in the collector current (I_C). This means that slight variations in the current flowing through the base can effectively control the current in the emitter-collector circuit.

Typically, Beta values range between 20 and 200 for most transistors. For instance, if a transistor has a Beta value of 100, one electron flows from the base terminal for every 100 electrons flowing between the emitter-collector terminals.

The mathematical relationship between Alpha (α) and Beta (β) expresses the current gain of the transistor. $\alpha = \frac{I_C}{I_E}$ and $\beta = \frac{I_C}{I_B}$ $\therefore I_C = \alpha I_E = \beta I_B$

$$As \quad \alpha = \frac{\beta}{\beta + 1} \quad and \quad \beta = \frac{\alpha}{1 - \alpha}$$

$$I_E = I_C + I_B \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 22.4$$

Where: "I_C" is the current flowing into the collector terminal, "I_B" is the current flowing into the base terminal and "I_E" is the current flowing out of the emitter terminal.

This type of bipolar transistor configuration has greater input impedance, current and power gain than that of the common base configuration but its voltage gain is much lower. The common emitter configuration is an inverting amplifier circuit. This means that the resulting output signal has a 180° phase-shift with regards to the input voltage signal.

Worked Example 22.1

Find Alpha (α) rating of the transistor shown in diagram. Hence determine the value of I_C using both α and β=49rating of the transistor.

Solution:

Step 1: Write down the known quantities and quantities to be found.

I_B=240μA, I_E=12mA and β=49

I_C=?

Step 2: Write down the formula and rearrange if necessary.

$$\alpha = \frac{\beta}{\beta + 1} = \frac{49}{49+1} = 0.98$$

The value of I_C can be found by using either α or β as under

$I_C = \alpha \times I_E$

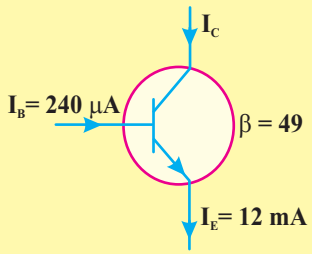
$I_C = 0.98 \times 12$ mA

I_C= 11.76 mA

$I_C = \beta \times I_B$

$I_C = 49 \times 240\mu A$

I_C= 11.76 mA



22.3.4 Operation Principles and I-V Characteristics and Methods of Transistor Biasing:

Bipolar Junction Transistor (BJT):

Operation Principles:

1. Emitter, Base, and Collector:

- The three layers of a BJT are called the emitter (E), base (B), and collector (C).
- The flow of current in an N-P-N transistor is from the emitter to the collector, and in a P-N-P transistor, it's from the collector to the emitter.

2. Transistor Action:

The operation of a BJT involves the injection of minority carriers (electrons in N-P-N or holes in P-N-P) from the emitter into the base region. This controls the majority carrier flow from the collector to the emitter.

I-V Characteristics:

Transistors use current-voltage characteristics to show how input voltage relates to output current. This helps understand how transistors amplify or switch signals.

Active Region:

Here, the transistor amplifies signals. A small current at the base-emitter junction (I_B) causes a larger current from collector to emitter (I_C). The graph shows a straight line, illustrating signal amplification.

Saturation and Cutoff:

Beyond the active region, these regions appear. In saturation, the transistor is ON, allowing maximum collector current. In cutoff, it's OFF, with minimal collector current.

Proper biasing keeps transistors in the active region for linear amplification. Understanding these characteristics is crucial for designing circuits for signal processing, amplification, and switching. Understanding these I-V characteristics is essential for designing circuits that utilize transistors for signal processing, amplification, and switching applications.

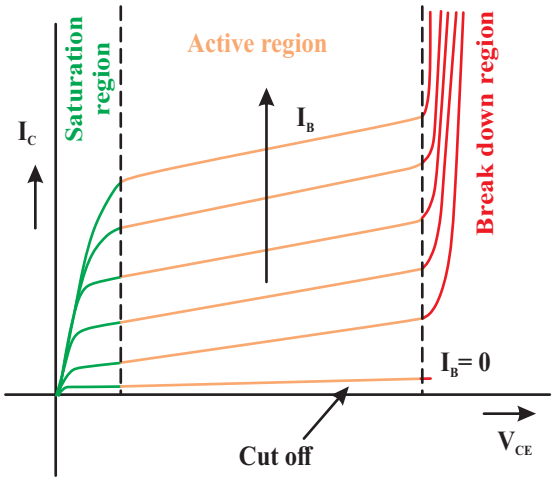


Figure 22.20. Active Region: Saturation and Cutoff

Biasing Methods:

Biasing is the process of applying external voltages to a transistor to establish a desired operating point for proper amplification or switching. There are several biasing methods for transistors. Here are some common biasing methods:

Bipolar Junction Transistor (BJT) Biasing Methods:

1. Fixed Bias (Base Bias):

- Connects a resistor between the base and the power supply to establish a fixed voltage at the base.
- Simple but not very stable.

2. Emitter Bias (Emitter Stabilized Bias):

- Connects a resistor from the emitter to the power supply, stabilizing the operating point.
- More stable than fixed bias.

3. Collector Feedback Bias:

- Combines features of fixed bias and emitter bias for improved stability.
- Uses a resistor network to provide feedback.

4. Voltage Divider Bias:

- Utilizes a resistor divider network to bias the base and provide stability.
- More stable than fixed bias but less stable than emitter bias.

22.3.5 The transistors as a switch:

A transistor can be used as a switch in electronic circuits, where it functions to either allow or block the flow of current. There are two main configurations for using a transistor as a switch: the NPN (negative-positive-negative) and PNP (positive-negative-positive) configurations.

NPN Transistor as a Switch:

The transistor Q_1 in figure 22.22 shows how to control output current with input.

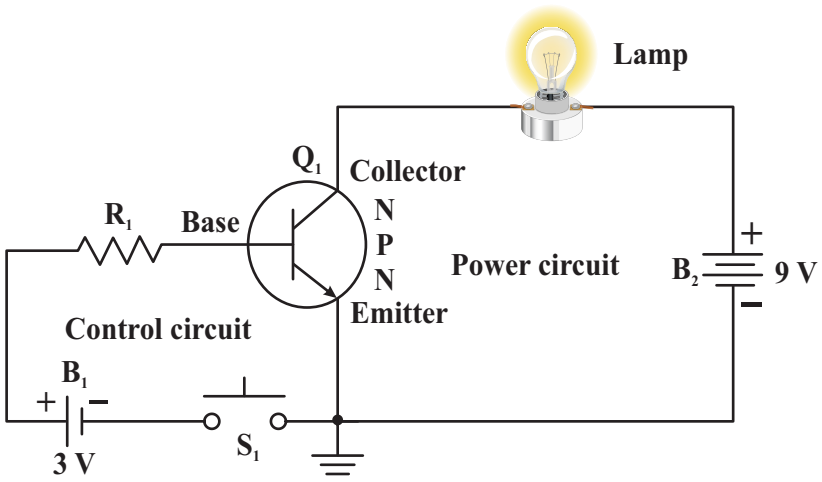


Figure 22.22 transistors as a switch

Here are key points to note:

1. Off State:

Normally, Q_1 allows no output current unless we apply forward voltage to its base-emitter circuit.

2. Forward Voltage:

The amount of output current is controlled by the forward voltage that controls base current. In figure 22.22:

- The input control circuit determines base current.
- The output current is collector current for the power circuit.
- Q_1 is an NPN transistor, needing a positive VBE for forward voltage.
- The emitter is common to both input control and power circuits.
- Common-emitter (CE) circuit is the most common transistor arrangement.

The base-emitter junction of Q_1 in figure 22.22 can be forward biased by battery B_1. Switch S_1 must be closed to apply forward voltage. Battery B_2 provides reverse voltage to the collector of Q_1. Reverse polarity means the collector is more positive than the base.

When switch S_1 is open:

- No current flows in the base-emitter or control circuit because no forward voltage is applied.
- Resistance from emitter to collector of the transistor is very high.
- No current flows in the power circuit, and the lamp doesn't light up.

When switch S_1 is closed:

- A small current flows in the control circuit.
- R_1 limits current in the base circuit.
- Resistance from emitter to collector of Q_1 decreases.
- A large current flows in the power circuit, lighting up the lamp.

Opening switch S_1 in the control circuit turns off the lamp in the power circuit because resistance from emitter to collector of Q_1 increases again, almost to infinity.

Transistor as an Amplifier:

Consider the circuit shown in figure 22.23. The dc operating voltages are shown in figure 22.23(a), without any input signal. The dc biasing levels are set by the feedback resistor R_2. The NPN transistor is biased so that the collector-to-emitter voltage Vc is half of the supply voltage. For the supply voltage of 10 v, therefore, the collector voltage is set at one-half the total, or +5 V. The 0.7 V at the base is partially turning on the transistor. The transistor acts as an amplifier when in this partially turned-on condition. It is the amount of dc forward bias that determines the operating level of the transistor.

(a)

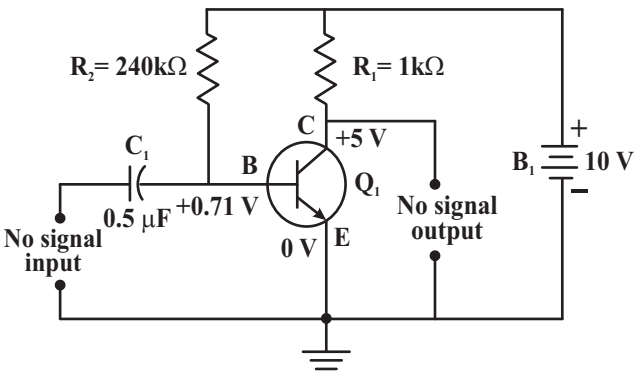


Figure 22.23 (a)
Transistors as an Amplifier

An input signal has been added to the amplifier in figure 22.23 (b), Input is coupled to the base by C_1. Amplified output is taken from the collector. The input signal is 0.02 V_{pp} as measured on an oscilloscope. The measured output signal voltage is 3 V_{pp}.The ac gain of the amplifier, therefore is calculated as

Av = Vout/ V= 3 /0.02 = 150

The output signal (3 V_{pp}) is 150 times greater than the input signal of 0.02 V_{pp}. This amplifier stage is said to have a voltage gain of 150.

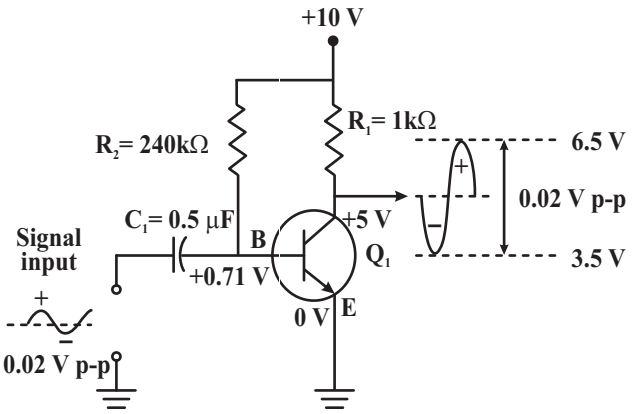


Figure 22.23 (a,b)
Transistors as an Amplifier

Worked Example 22.2

In a common base connection, I_C = 0.95 mA and I_B = 0.05 mA. Find the value of α.

Solution:

Step 1: Write down the known quantities and quantities to be found.

I=0.95mA I_B = 0.05mA R = 470 Ohm α.= ?

Step 2: Write down the formula and rearrange if necessary.

We know $I_E = I_B + I_C$ = 0.05+0.95 =1mA

Current amplification factor = $\propto = \frac{I_C}{I_E} = \frac{0.95}{1} = 0.95$

Result: 0.95

22.3.6 Common Collector (CC) Circuit:

In CC configuration, the input signal is applied to the base, and the output signal is taken from the emitter, with the collector terminal serving as the common connection.

Figure 22.24 depicts this setup. Here, when the base-emitter junction is forward biased, a small base current I_B causes a much larger collector current I_C to flow. The emitter current I_E is approximately equal to I_C since $I_E = I_B + I_C$ and I_B is significantly smaller than I_C.

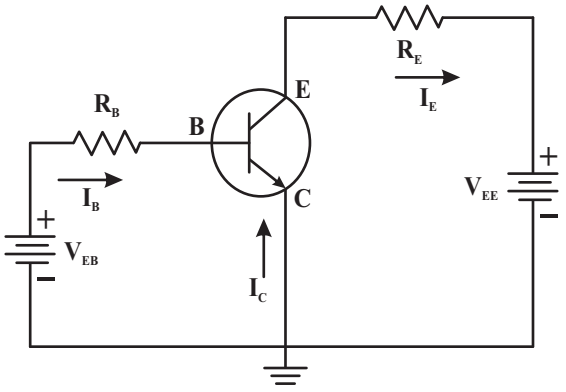


Figure 22.24 common collector circuits

Advantages:

CC amplifiers offer high current gain and low input impedance, making them suitable for applications where a high-resistance input needs to drive a low-resistance output load.

Applications:

Commonly used in impedance matching and signal buffering circuits due to their favorable characteristics.

Common Base (CB) Circuit:

In CB configuration, the input and output signals share the base terminal of the transistor.

Figure 22.25 demonstrates this configuration. Here, the input signal is applied to the emitter, and the output signal is taken from the collector. The base-emitter junction is forward biased to allow transistor operation.

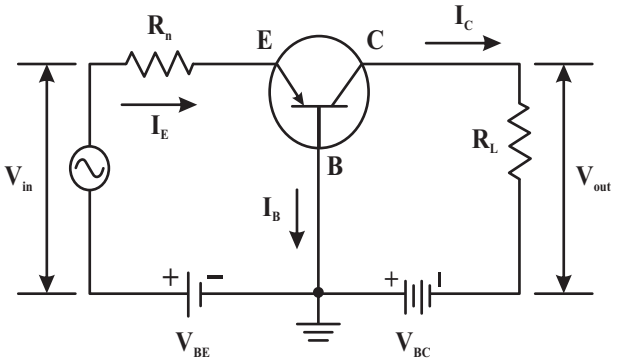


Figure 22.5 Common Base (CB) Circuit

Characteristics:

CB amplifiers provide low input impedance and approximately unity voltage gain. The current gain in CB configuration is expressed as $\frac{I_{out}}{I_{in}}$ or the formula $\frac{I_c}{I_E}$

However, as the base current is extremely small compared to the collector current, the emitter current is therefore approximately equal to the collector current. Thus $I_E \approx I_C$

Applications:

Less commonly used compared to CC and Common Emitter (CE) configurations, CB amplifiers are employed in specific applications where low input impedance and unity voltage gain are advantageous.

22.4 Operational Amplifier:

An operational-amplifier is a **direct coupled high gain amplifier**. It can be operated on both with AC and DC signals. Operational Amplifier, also called as an Op-Amp, is an integrated circuit, which can be used to perform various linear, non-linear, and mathematical operations as shown in figure 22.26.

Op-amps are ideal for DC amplification and are used often in signal conditioning, filtering or other mathematical operations (add, subtract, integration and differentiation).

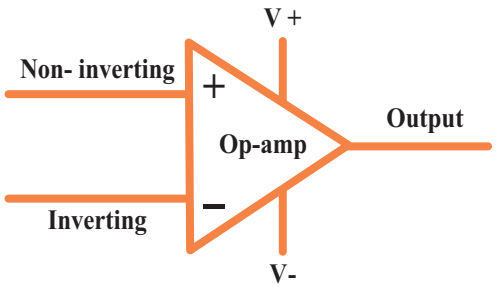


Figure 22.26 Operational Amplifier

22.4.1 Properties of an Ideal Operational Amplifier (Op-Amp):

The ideal op-amp has the following properties:

Characteristics	Ideal Op-Amp
Infinite open-loop voltage gain	An ideal op amp is a device often used as an amplifier. When you input voltage into it, the op amp outputs a amplified voltage. In an ideal scenario, the op amp would provide extremely high gain, essentially infinite, amplifying the signal countless times for maximum gain as needed.
Infinite input resistance (Impedance)	An ideal op amp has super high input impedance, which means it won't load the circuit. If input impedance is low, the op amp draws more current; if it's high, less current is drawn. We aim for high input impedance to avoid disturbing the original circuit by minimizing current pulled from it, ideally with infinite input impedance.
Zero output resistance (Impedance)	In an ideal op amp, the output impedance is zero. This means when the op amp generates a signal, we want it to have zero resistance, ensuring that the maximum voltage goes to the output load. In a circuit, voltage gets divided based on the impedance. Higher impedance means more voltage drop. For the voltage to drop across the output load, the load's impedance should be greater than the op amp's output. That's why, ideally, we aim for zero output impedance in the op amp.
Gain independent of frequency	In an ideal op amp, the produced gain remains constant regardless of the input signal's frequency. This means the amplification stays reliable and consistent across all frequencies.
Infinite bandwidth	The bandwidth of an op-amp is the range of frequencies it amplifies equally. An ideal op-amp amplifies all frequencies, so it should ideally have an infinite bandwidth.
Infinite slew rate	An ideal op-amp should change the output instantaneously as the input is changed. The slew rate of the op-amp is the factor that affects this time delay. An infinite slew rate means there is no time delay.
Zero noise contribution	Any signal includes a small amount of noise. The ideal op-amp does not produce any noise itself, although it will amplify any noise that is present in its input.

22.4.2 Op-Amp as comparator:

The op-amp shown in figure 22.27 is connected to two power supplies. One battery of 9V is connected between the zero-volt line and the +9V positive supply terminal of the op-amp and the other between the zero-volt line and the negative power supply terminal. These batteries are not shown.

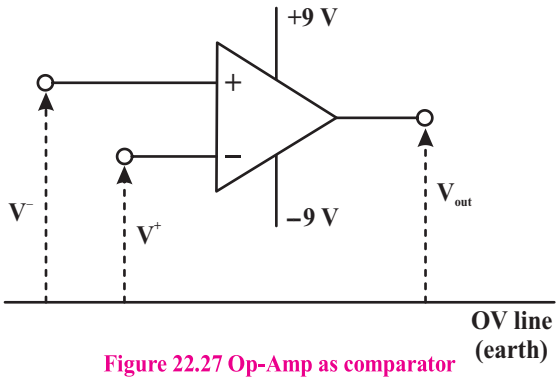


Figure 22.27 Op-Amp as comparator

The output voltage is given by $V_{out} = G_o \times (V^+ - V^-)$, where G_o is the open-loop voltage gain. Notice that all of these voltages are measured with reference to the zero-volt line, which is often connected to earth.

Suppose that $G_o = 10^5$ and that $V^+ = 0.15$ V and $V^- = 0.10$ V. Then equation becomes

$V_{out} = 10^5 \times (0.15 - 0.10) = 5000$ V

Certainly, it's impossible. The op-amp is therefore saturated and V_{out} will be close to one of the power supply voltages, in this case 9 V because V^+ is bigger than V^-.

In fact:

- If V^+ is slightly greater in magnitude than V^-, then V_{out} will have magnitude equal to the positive power supply voltage.
- If V^+ is slightly smaller in magnitude than V^-, then V_{out} will have magnitude equal to the negative power supply voltage.

The op-amp is serving as a comparator, comparing V^+ and V^-. If these voltages are slightly different, then the output voltages tell us which one is larger.

22.4.3 Effects of negative feedback on the gain of an operational amplifier:

Negative Feedback:

Negative feedback occurs when a portion of the output signal, like voltage or current, is feedback into the input in a way that opposes or subtracts from the input signal.

Advantages:

1. Reduced Gain:
   Negative feedback decreases the overall gain of the system.
2. Decreased Distortion and Noise:
   It minimizes distortion and noise, leading to a cleaner output signal.
3. Improved Stability:
   The gain becomes more stable and less sensitive to external changes, such as temperature.
4. Increased Bandwidth:
   Enhances the system's bandwidth.
5. Better Impedance:
   Lowers the output impedance and raises the input impedance.

Operation in Electronic Systems:

Feedback is unidirectional, flowing from output to input, ensuring the loop gain (G) remains independent of load and source impedances. A summing point at the input subtracts the feedback signal from the input signal, creating an error signal that drives the system.

Example Circuit (Figure 2.28):

Setup:

Input $V_{in}$is connected to the non-inverting input (+) of an op-amp. The output $V_{out}$is fed back to the inverting input (-).

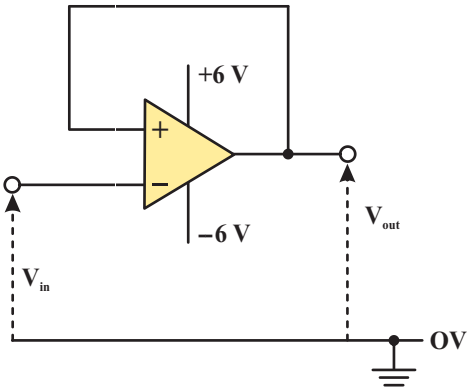


Figure 2.28 Operations in Electronic Systems

Operation:

With an infinite open-loop gain and no saturation, V^- equals V^+.As V_{in} changes, V_{out} adjusts to maintain , $V^- = V^+$.Equation: $V_{out} = G_o \times (V^+ - V^-)$

Simplified: Since$V_{out} = V^-$ and $V_{in} = V^+$ then we have

$V_{out} = G_o (V_{in} - V_{out})$

$V_{out} (1+G_o) = G_o V_{in}$

The closed-loop gain G is given by: $G = \frac{V_{out}}{V_{in}} = \frac{G_o}{1+G_o}$ 22.5

As G_o is very high about 10^5, there is little difference between G_o and $(1+G_o)$, so the closed-loop gain is very nearly 1. This analysis is true as long as the output voltage is smaller than the supply voltage, in this case as long as V_{out} is between -6 v and +6 v.

Practical Example:

A Piezo-electric microphone with high internal resistance can be connected to an op-amp to increase current output while maintaining the same voltage.

Negative Feedback Benefits Summary:

- Less distortion.
- Increased bandwidth.
- The gain is more stable and not affected by changes in temperature, etc.

The output resistance (impedance) can be low and the input resistance high.

22.4.4 The circuit diagrams for both the inverting and the non-inverting amplifier for single signal input:

Inverting operational amplifier:

In inverting operational amplifiers, the op amp forces the negative terminal to equal the positive terminal, which is commonly ground. Therefore, the input current is determined by the V_{IN} / R_1 ratio as shown in Figure 22.29: Inverting Operational Amplifier.

In this configuration, the same current flows through R_2 to the output. Ideally, current does not flow

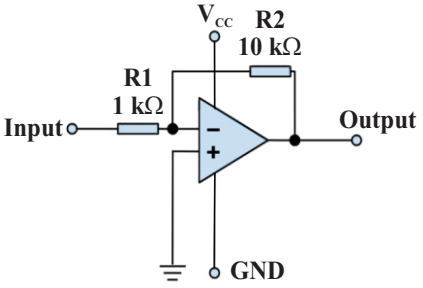


Figure 22.29
circuit diagrams for inverting Amp

into the operational amplifier's negative terminal due to its high Z_{IN}. The current flowing from the negative terminal through R_2 creates an inverted voltage polarity with respect to V_{IN}. This is why these op amps are labeled with an inverting configuration. Note that the op amp's output can only swing between its positive and negative supplies, so creating a negative output voltage requires an op amp with a negative supply rail. V_{OUT} can be calculated with Equation

$$V_{out} = -\frac{R_2}{R_1} \times V_{in} \ldots\ldots\ldots\ldots\ldots\ldots 22.6$$

Non-inverting operational amplifier:

In a non-inverting amplifier circuit, the input signal from the source is connected to the non-inverting (+) terminal as shown in figure 22.30. Non-Inverting Operational Amplifier

The operational amplifier forces the inverting (-) terminal voltage to equal the input voltage, which creates a current flow through the feedback resistors. The output voltage is always in phase with the input voltage, which is why this topology is known as non-inverting. Note that with a non-inverting amplifier, the voltage gain is always greater than 1, which is not always the case with the inverting configurations. V_{out} can be calculated with Equation:

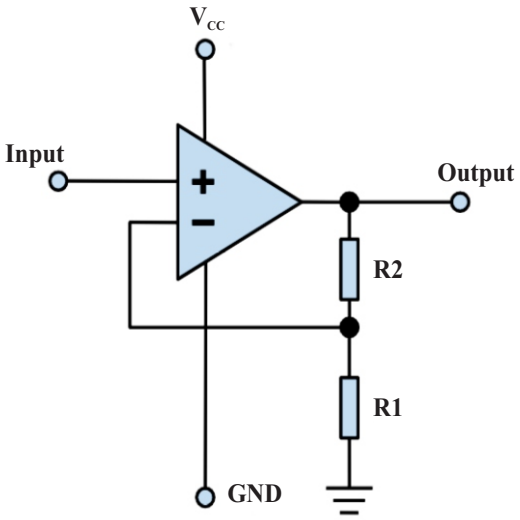


Figure 22.30 circuits for non-inverting amplifier

$$Vout = \frac{1 + R_2}{R_1} \times Vin \ldots\ldots\ldots\ldots\ldots\ldots 22.7$$

22.4.5 Virtual Earth Approximation:

To understand how the inverting amplifier works, you need to understand the concept of virtual earth approximation. In this approximation the potential at the inverting input (-) is very close to 0 V. There are two steps in the arguments:

The op-amps multiplies the difference in potential between the inverting and non-inverting inputs, V^- and V^+, to produce the output voltage V_{out}. Because the open-loop voltage gain is very high, the difference between V^- and V^+ must be almost zero.

The non-inverting input (+) is connected to the zero-volt line so $V^+ = 0$. Thus V^- must be close to zero and the inverting input (-) is almost at earth potential.

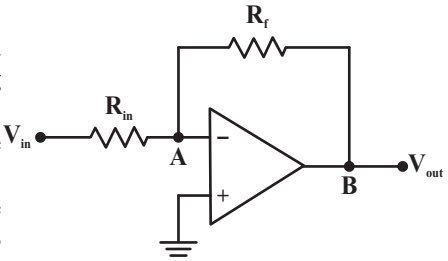


Figure 22.31 virtual earth approximation

The point Ais known as a virtual earth as shown in figure 22.31. It cannot actually be 0V but it is very close to 0V. This approximation is true as long as the op-amp is not saturated, and for frequencies where the open-loop voltage gains is high.

Applications:

This approximation is often used in summing amplifier circuits and other applications where it simplifies analysis.

Summing Amplifier Example:

- In a summing amplifier, multiple input voltages can be summed using resistors connected to the inverting input.
- The virtual ground assumption makes it easier to analyze the currents and voltages in the circuit.

Gain of an Inverting Amplifier:

If the current in the input resistor R_{in} is I_{in} and the current in the feedback resistor R_f is I_f, then point P is at 0V:

$$I_{in} = \frac{V_{in}}{R_{in}} \quad and \quad I_f = \frac{V_{out}}{R_f}$$

The input resistance of the op-amp is very high and so virtually no current enters or leaves the inverting input (-) of the op-amp. This means that I_{in} and I_f must be equal in size.
If V_{in} is apositive potential, then the current in two resistors flows from left to right. V_{out} will be negative because the current flows from P, which is at 0 V, to the output connection, which must have a lower voltage than 0 V. Thus

$$I_f = -I_{in} \quad \text{and} \quad \frac{V_{out}}{Rf} = -\frac{V_{in}}{R_{in}}$$

The gain of the inverting amplifier is thus given by

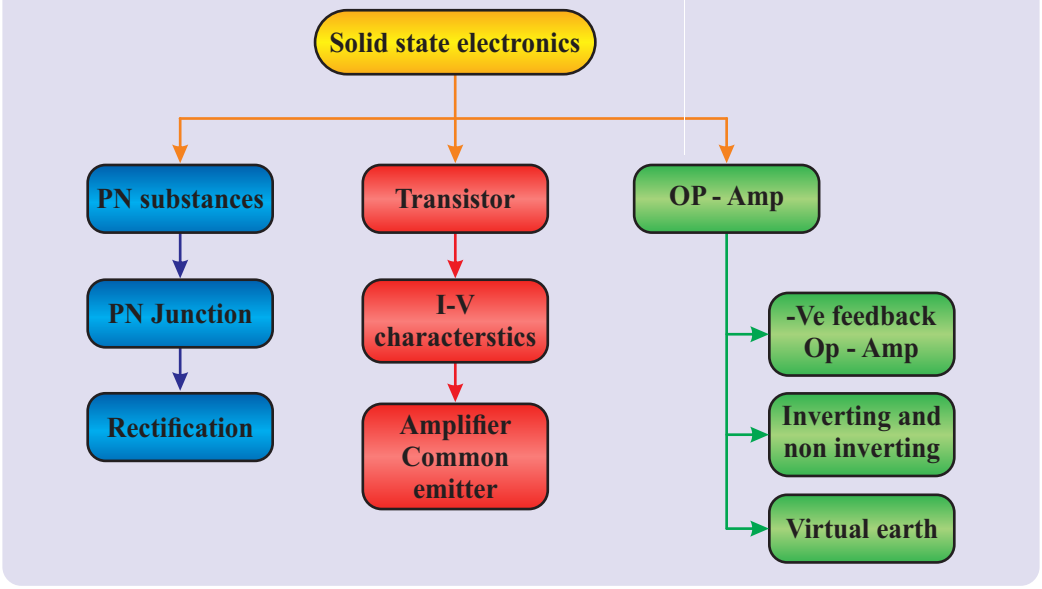
$$G = \frac{V_{out}}{V_{in}} = -\frac{R_f}{R_{in}} \ldots\ldots\ldots 22.8$$

The negative sign shows that when the input voltage is positive the output voltage is negative and when the input is positive the output is negative. If the input voltage is alternating then there will be a phase difference of 180^o or π rad between the input and the output voltages.



- ✓ A *p*-type substance is a semiconductor material which has an excess of positively charged particles called "holes" as its majority charge carriers.
- ✓ An *n*-type substance is a semiconductor material that has an excess of negatively charged electrons as its majority charge carriers.
- ✓ An intrinsic semiconductor is a pure semiconductor material with no intentional impurities added.
- ✓ Doping is the process of intentionally introducing impurities into a semiconductor crystal to modify its electrical properties.
- ✓ Doping with Donor Impurities: A small amount of a group V element is introduced into the crystal lattice of the semiconductor. Common donor impurities include phosphorus (P) or arsenic (As).
- ✓ Doping with Acceptor Impurities: A small amount of a group III element is introduced into the crystal lattice of the semiconductor. Common acceptor impurities include boron (B) or gallium (Ga).
- ✓ P-N Junction is formed when the *p*-type and *n*-type semiconductors are joined, is called as *p-n* junction.
- ✓ P-N Junction Diode is a semiconductor device that is formed through this method and used in allowing the flow of electric current in one direction and blocking in the opposite.
- ✓ Forward Bias: The *P-N* junction is forward-biased when the *P*-type is connected to the positive terminal of the battery and the *N*-type is connected to the negative terminal.
- ✓ Reverse Bias: The *P-N* junction is reverse biased when the*P*-type is connected to the negative terminal of the battery and the *N*-type is connected to the positive side.
- ✓ Rectification is the process of converting an alternating current (AC) waveform into a direct current (DC) waveform, i.e., creating a new waveform that has only a single polarity.
- ✓ A light -emitting diode (LED) is a *P-N* junction in which current flows as it becomes forward biased.
- ✓ A photodiode is one type of light detector which uses to convert the light into current or voltage based on the mode of operation of the device.
- ✓ Photo Voltaic Cells or Solar cells, also called photovoltaic cells, are rather heavily doped *p-n* junction diodes used to convert sunlight into electric energy.
- ✓ Transistor: The Bipolar Junction Transistor was invented in 1948 by J. Bardeen, W. Shockley and W. Brattain.
- ✓ It consists of a crystal of one type of doped semiconductor sandwiched between two of the opposite type.

CONCEPT MAP



EXERCISE

Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. A semiconductor is an element with a valence electron _____.
   (a) Four (b) Eight (c) Two (d) One
2. A pure semiconductor is known as
   (a) Extrinsic (b) Intrinsic (c) Transistor (d) Diode
3. An acceptor atom is also called
   (a) Penta-valent atom (b) Trivalent atom
   (c) Minority carrier (d) Majority carrier
4. With which one of the following elements silicon should be doped so as to give p-type of semiconductor?
   (a) Germanium (b) Arsenic (c) Selenium (d) Boron
5. For a full-wave rectifier, the output frequency
   (a) Equals one-half the input frequency (b) Equals the line frequency
   (c) Equals two times the input (d) Is three times the line frequency

6. LED construction needs a semiconductor material is:
   (a) Silicon (b) Germanium (c) Gallium (d) Gallium arsenide
7. The frequency of a half-wave signal is
   (a) Twice the line frequency (b) Equal to the line frequency
   (c) One-half of the line frequency (d) One-fourth the line frequency
8. The voltage gain of an emitter follower circuit is
   (a) High (b) Low (c) Very high (d) Moderate
9. What is also called as the conventional amplifier?
   (a) Common-collector circuit (b) Emitter follower circuit
   (c) Common base circuit (d) Common emitter circuit
10. An op-amp with negative feedback provides _____ output parameter.
   (a) Gain (b) Bandwidth
   (c) Input-output impedance (d) Amplified

Section (B): CRQs (Short Answered Questions):

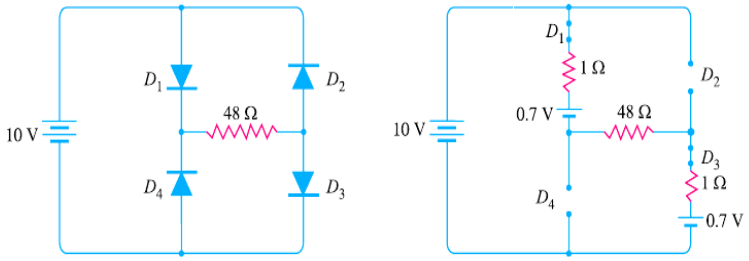
1. Describe a *p-n* junction (diode) and how holes and electrons are produced in semiconductor?
2. Define and distinguish between p-n-p & n-p-n transistors?
3. Explain common-base and common collector configurations?
4. Describe the operations of transistors.
5. Explain the use of transistors as a switch and an amplifier (common-emitter).
6. How would you understand the effects of negative feedback on the gain of an operational amplifier?
7. Draw and briefly describe the circuit diagrams for both the inverting and the non-inverting amplifier for single signal input?
8. Derive an expression for the gain of inverting amplifiers by using virtual earth approximation.
9. Describe the properties of an ideal operational amplifier.

Section (C): ERQs (Long Answered Questions):

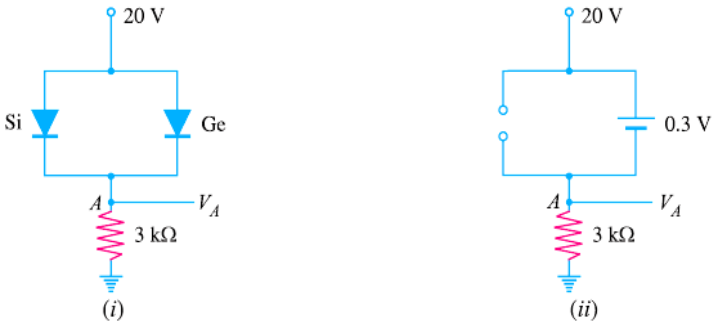
1. Define Intrinsic (pure) and doped semiconductors. And How the N-type and P-type semiconductors are produced
2. Describe a P-N junction (diode) discus its forward and reverse biasing.
3. Describe the I-V characteristic curves of P-N junction.
4. Define rectification and describe the use of diodes for half and full wave rectifications.
5. Describe the function and use of LED, Photodiode and Photo voltaic cell Describe the function and use of LED, Photodiode and Photo voltaic cell.
6. Explain the use of transistors as a switch and an amplifier (common-emitter).
7. Describe the properties of an ideal operational amplifier.

Section (D): Numerical:

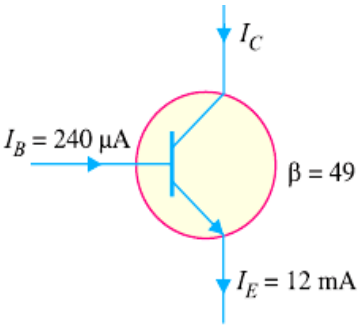
1. A Ge diode has a voltage drop of 04 V when 12 mA flow through it. If the same 470 Ohm is used, what battery voltage is needed? **(6.04 V)**
2. A semiconductor diode laser has a peak emission wavelength of 1.55 μm. Find its band gap in eV. **(0.8 eV)**
3. Calculate the current through 48 Ω resistor in the circuit shown in Fig. (i). Assume the diodes to be of silicon and forward resistance of each diode is 1 Ω. **(8.6 V, 50 Ohm, 172 mA)**



4. Find the voltage V_A in the circuit shown in Fig.(i). Use simplified model. (19.7 V)

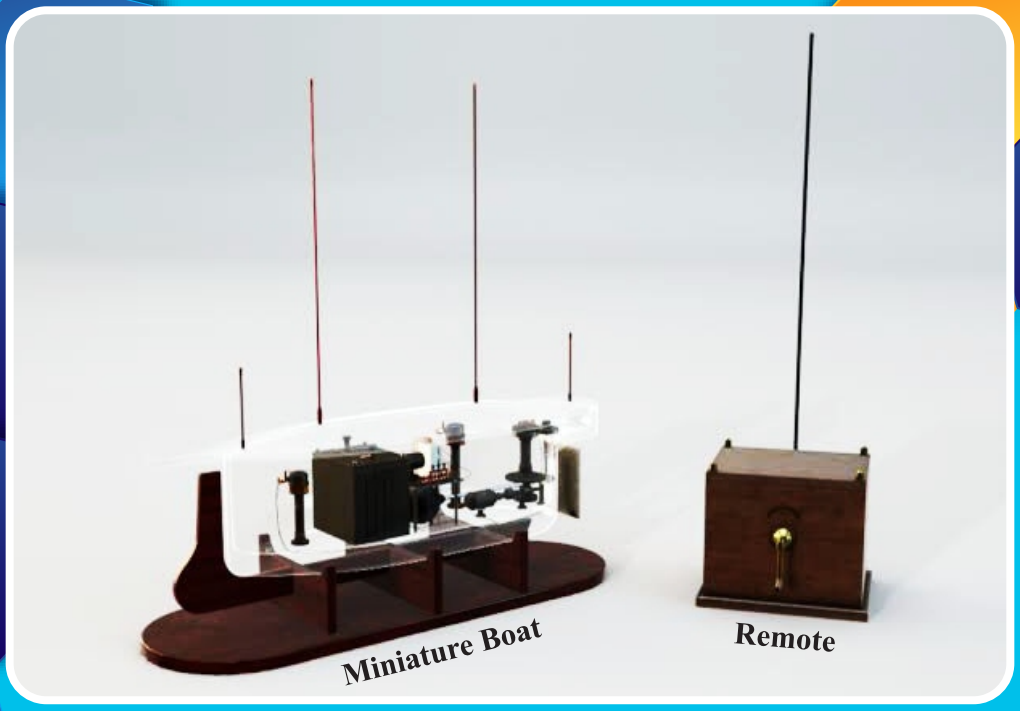


5. In a common base connection, I_E = 1mA, I_C = 0.95mA. Calculate the value of I_B. **(0.05 mA)**
6. Find the value of β if (i) α = 0.9 (ii) α = 0.98 (iii) α = 0.99. **(9, 49, 99)**
7. Calculate I_E in a transistor for which β = 50 and I_B = 20 μA. **(1.02 mA)**
8. Find the α rating of the transistor shown in Fig. Hence determine the value of I_C using both α and β rating of the transistor. **(0.98, 11.76)**



Unit 23 Digital Electronics

Teaching Periods 03 **Weightage %** 02



1st Remote of world , In 1898, Nikola Tesla invented "teleautomation", which he demonstrated with a radio-controlled miniature boat, the first remote control device ever made. The remote control system contained a method for decoding Hertzian waves. This system would toggle actions based on different signals, functioning as an AND gate.

In this unit student should be able to:

- Explain signal levels employed in digital electronics or circuits are 'high' (for example, +5 volts) and 'low' (for example, 0 volts)
- Describe the 'high' and 'low' states are referred to as '1' and '0' for open and close circuit respectively
- Recall electronic symbols of the logic gates AND, OR, NOT, NOR, NAND, XOR
- Use logic gates (AND, OR, NOT, NOR, NAND, XOR, combinations) and their truth tables for 2, 3 and 4 inputs
- Identify the behavior of a 2 inputs AND gate and a 2 inputs OR gate with simple circuit using 2 switches, a lamp and a battery

Introduction:

Digital electronics is a branch of electronics that deals with the manipulation of digital signals, represented as binary digits (0s and 1s).

Unlike analog electronics, which uses continuous signals, digital electronics processes discrete signals, enabling precise control and storage of information. This field encompasses the design, analysis, and application of digital circuits and systems, forming the foundation for modern computing, communication, and information processing technologies.

23.1 Digital signal levels:

In digital electronics, signal levels are represented in various ways. Here are common digital signal levels:

1. Low Level (0):

- This represents the binary digit 0.
- In terms of voltage, it is associated with a lower voltage level.
- Often referred to as "low" or "logic 0."

2. High Level (1):

- This represents the binary digit 1.
- In terms of voltage, it is associated with a higher voltage level.
- Often referred to as "high" or "logic 1."

3. Threshold Level:

- The threshold level is the voltage level that separates low and high states.
- Signals below this threshold are interpreted as 0, and signals above it are interpreted as 1.
- The threshold level helps define the noise margin and ensure reliable signal interpretation.

4. Logic Levels:

- Depending on the technology and the system, different logic levels might be used.

5. Swing or Voltage Range:

- The difference in voltage levels between low and high states is often referred to as the swing or voltage range. For example, voltage varies from -5V to +5V will have a swing or voltage range of 10V.
- A larger voltage swing can enhance noise immunity and signal reliability.

DO YOU KNOW?

Noise immunity is a measure of how well a logic circuit can withstand unwanted electrical signals or noise without reacting to them.

DO YOU KNOW?

If we deal with mathematics, then the binary system, using 1 or 0, will be employed. If we are dealing with voltage in a circuit, then HIGH or LOW can be utilized.

23.1.2 Understanding Digital Signal Levels:

In digital electronics, the 'high' and 'low' states are often symbolically represented as '1' and '0,' respectively. This binary notation helps describe the states of open and closed circuits.

- **High State ('1'):** When a digital signal is in a 'high' state, it is typically associated with a closed circuit or a voltage level close to the maximum specified value. The symbolic representation for this state is '1'. It signifies the 'on' or 'active' state in many digital systems.
- **Low State ('0'):** Conversely, the 'low' state is linked to an open circuit or a voltage level close to the minimum specified value. It is symbolically represented as '0'. This state signifies the 'off' or 'inactive' state in digital systems.

High and Low represent the **signal levels** in the circuit. Digital waveforms consist of voltage levels that are changing back and forth between the HIGH and LOW states. These can also be described as the one that is composed of series of pulses, or a pulse train as shown in Figure 23.1.

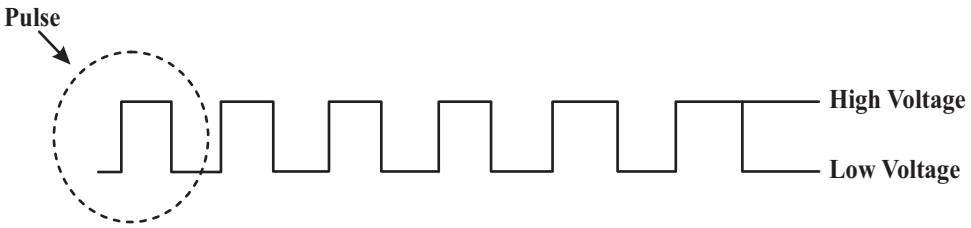


Figure 23. 1 Pulse showing High and Low states

These symbolic representations simplify the expression and interpretation of digital information, allowing for clear communication and logical operations within electronic circuits. The '1' and '0' states form the basic building blocks for encoding and processing binary data in digital systems. This binary representation forms the fundamental language of digital systems, allowing them to process and manipulate data efficiently.

DO YOU KNOW?

The mathematics of binary system is different from ordinary mathematics. The special algebra representing binary system is known as Boolean Algebra.

23.2 Logic gates:

Logic gates are tiny electronic switches that enable decision-making in computers and other devices. They combine to perform logical operations, playing a crucial role in information processing.

23.2.1 Types of Logic Gates:

Logic gates are the basic components of the digital circuit with one output and more than one input. AND, OR, and NOT gates are the basic gates, while NAND and NOR are the universal gates. EX-OR and EX-NOR are the special gates.

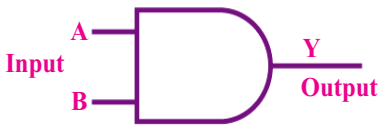


Figure 23. 2 AND gate

1. AND Gate:

The AND gate produces a 'high' output only when all its inputs are 'high.' In other words, the output of an AND gate attains state 1 if and only if all the inputs are in state 1.

The Boolean expression of AND gate and the Truth Table No:23.1 is given as for

1. Two-input Y = A.B
2. Three-input Y = A.B.C
3. Four input Y = A.B.C.D

Truth Table No. 23.1: AND

Two Input *Y = A.B*			**Three Input** *Y = A.B.C*				**Four Input** *Y = A.B.C.D*				
A	***B***	***Y***	***A***	***B***	***C***	Y	A	B	C	D	Y
0	0	0	0	0	0	0	0	0	0	0	0
0	1	0	0	0	1	0	0	0	0	1	0
1	0	0	0	1	0	0	0	0	1	0	0
1	1	1	0	1	1	0	0	0	1	1	0
			1	0	0	0	0	1	0	0	0
			1	0	1	0	0	1	0	1	0
			1	1	0	0	0	1	1	0	0
			1	1	1	1	0	1	1	1	0
							1	0	0	0	0
							1	0	0	1	0
							1	0	1	0	0
							1	0	1	1	0
							1	1	0	0	0
							1	1	0	1	0
							1	1	1	0	0
							1	1	1	1	1

2. OR Gate:

An OR gate generates a 'high' output if any of its inputs are 'high'. In other words, the output of an OR gate attains state 1 if one or more inputs attain state 1.

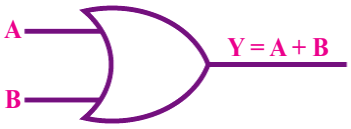


Figure 23. 3 OR gate

The Boolean expression of OR gate and the Truth Table No:23.2 is given as for

1. Two-input Y = A+B
2. Three-input Y = A+B+C
3. Four input Y = A+B+C+D

Truth Table No. 23.2: OR Gate

Two Input *Y = A+B*			Three Input *Y = A+B+C*				Four Input *Y = A+B+C+D*				
A	***B***	***Y***	***A***	***B***	***C***	Y	A	B	C	D	Y
0	0	0	0	0	0	0	0	0	0	0	0
0	1	1	0	0	1	1	0	0	0	1	1
1	0	1	0	1	0	1	0	0	1	0	1
1	1	1	0	1	1	1	0	0	1	1	1
			1	0	0	1	0	1	0	0	1
			1	0	1	1	0	1	0	1	1
			1	1	0	1	0	1	1	0	1
			1	1	1	1	0	1	1	1	1
							1	0	0	0	1
							1	0	0	1	1
							1	0	1	0	1
							1	0	1	1	1
							1	1	0	0	1
							1	1	0	1	1
							1	1	1	0	1
							1	1	1	1	1

3. NOT Gate:

It is used to perform the inversion operation in digital circuits, For example, the output of a NOT gate attains state 1 if and only if the input does not attain state 1.



Figure 23. 4 : NOT gate

The Boolean expression is: $Y = \overline{A}$

It is read as Y equals NOT A.

The truth table of NOT gate is given in table 23.3.

Truth Table No. 23.3 NOT GATE

A	Y
0	1
1	0

4. NAND Gate:

Similarly, a NAND gate is an AND gate followed by a NOT gate, yielding the opposite output ofan AND gate.

The Boolean expression of NAND gate and the Truth Table No:23.4 is given as for

1. Two-input $Y = \overline{A.B}$
2. Three_input $Y = \overline{A.B.C}$
3. Four input $Y = \overline{A.B.C.D}$

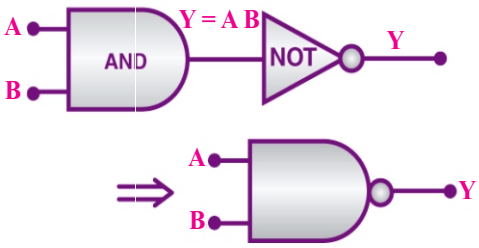


Figure 23.5 NAND GATE

Truth Table No. 23.4: NAND Gate

Two Input $Y = \overline{A.B}$			Three Input $Y = \overline{A.B.C}$				Four Input $Y = \overline{A.B.C.D}$				
A	***B***	***Y***	***A***	***B***	***C***	Y	A	B	C	D	Y
0	0	1	0	0	0	1	0	0	0	0	1
0	1	1	0	0	1	1	0	0	0	1	1
1	0	1	0	1	0	1	0	0	1	0	1
1	1	0	0	1	1	1	0	0	1	1	1
			1	0	0	1	0	1	0	0	1
			1	0	1	1	0	1	0	1	1
			1	1	0	1	0	1	1	0	1
			1	1	1	0	0	1	1	1	1
							1	0	0	0	1
							1	0	0	1	1
							1	0	1	0	1
							1	0	1	1	1
							1	1	0	0	1
							1	1	0	1	1
							1	1	1	0	1
							1	1	1	1	0

5. NOR Gate:

A NOR gate is an OR gate followed by a NOT gate, providing the opposite result of an OR gate.

The Boolean expression of NOR gate and the Truth Table No:23.5 is given as for

1. Two-input $Y = \overline{A + B}$
2. Three-input $Y = \overline{A + B + C}$
3. Four input $Y = \overline{A + B + C + D}$

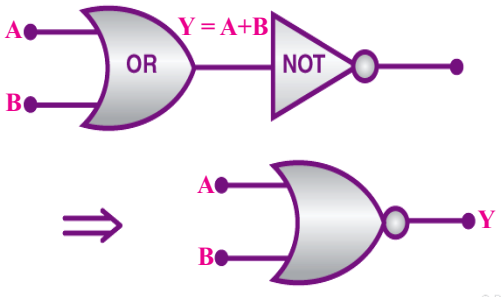


Figure 23. 6
NOR gate and its equivalent diagram

Truth Table No. 23.5: NOR Gate

Two Input $Y = \overline{A + B}$			Three Input $Y = \overline{A + B + C}$				Four Input $Y = \overline{A + B + C + D}$				
A	***B***	***Y***	***A***	***B***	***C***	Y	A	B	C	D	Y
0	0	1	0	0	0	1	0	0	0	0	1
0	1	0	0	0	1	0	0	0	0	1	0
1	0	0	0	1	0	0	0	0	1	0	0
1	1	0	0	1	1	0	0	0	1	1	0
			1	0	0	0	0	1	0	0	0
			1	0	1	0	0	1	0	1	0
			1	1	0	0	0	1	1	0	0
			1	1	1	0	0	1	1	1	0
							1	0	0	0	0
							1	0	0	1	0
							1	0	1	0	0
							1	0	1	1	0
							1	1	0	0	0
							1	1	0	1	0
							1	1	1	0	0
							1	1	1	1	0

6. XOR Gate (Exclusive-OR Gate):

The XOR gate outputs a 'high' signal if the number of 'high' inputs is odd. In other words, the output of a two-input XOR gate attains state 1 if one adds only input and attains state 1.

The Boolean expression of XOR gate and Truth Table No:23.6 is given as for

1. $Y = A.\overline{B} + \overline{A}$
2. Three-input
   $Y = A.\overline{B}.\overline{C} + \overline{A}.B.\overline{C} + \overline{A}.\overline{B}.C + A.B.C$
3. Four input
   $Y = A \oplus B \oplus C \oplus D$

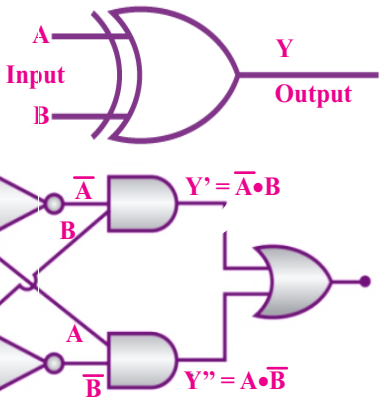


Figure 23. 7 XOR gate and its equivalent diagram

Truth Table No. 23.6: NOR Gate

Two Input $Y = A.\overline{B} + \overline{A}$			Three Input $Y = A.\overline{B}.\overline{C} + \overline{A}.B.\overline{C} + \overline{A}.\overline{B}.C + A.B.C$				Four Input $Y = A \oplus B \oplus C \oplus D$				
A	***B***	***Y***	***A***	***B***	***C***	Y	A	B	C	D	Y
0	0	0	0	0	0	0	0	0	0	0	0
0	1	1	0	0	1	1	0	0	0	1	1
1	0	1	0	1	0	1	0	0	1	0	1
1	1	0	0	1	1	0	0	0	1	1	0
			1	0	0	1	0	1	0	0	1
			1	0	1	0	0	1	0	1	0
			1	1	0	0	0	1	1	0	0
			1	1	1	1	0	1	1	1	1
							1	0	0	0	1
							1	0	0	1	0
							1	0	1	0	0
							1	0	1	1	1
							1	1	0	0	0
							1	1	0	1	1
							1	1	1	0	1
							1	1	1	1	0

7. XNOR Gate (Exclusive-NOR Gate):

In the XNOR gate, the output is in state 1 when both inputs are the same, that is, both 0 or both 1.

The Boolean expression of XNOR gate and the Truth Table No:23.7 is given as for

1. $Y = ((A \oplus \overline{B})) = (A.B + \overline{A}.\overline{B})$
2. Three-input
   $Y = ((\overline{A}).(\overline{B})\,\overline{C}) + (A.B.C)$
3. Four input
   $Y = (\overline{A}\,.\,(\overline{B}.)\,\overline{C}\,.\,\overline{D}) + (A.B.C.D)$

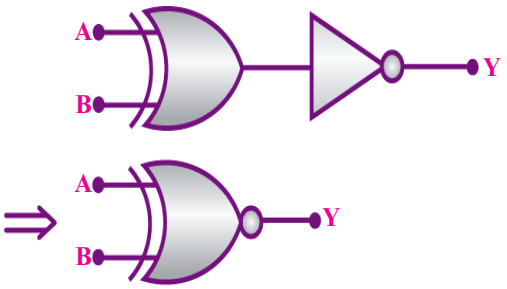


Figure 23.8: XOR gate and its equivalent diagram

Truth Table No. 23.7: XNOR Gate

Two Input $Y = ((A \oplus \overline{B})) = (A.B + \overline{A}.\overline{B})$			**Three Input** $Y = ((\overline{A}).(\overline{B})\,\overline{C}) + (A.B.C)$				**Four Input** $Y = (\overline{A}\,.\,(\overline{B}.)\,\overline{C}\,.\,\overline{D}) + (A.B.C.D)$				
A	***B***	***Y***	***A***	***B***	***C***	Y	A	B	C	D	Y
0	0	1	0	0	0	1	0	0	0	0	1
0	1	0	0	0	1	0	0	0	0	1	0
1	0	0	0	1	0	0	0	0	1	0	0
1	1	1	0	1	1	1	0	0	1	1	1
			1	0	0	0	0	1	0	0	0
			1	0	1	1	0	1	0	1	1
			1	1	0	1	0	1	1	0	0
			1	1	1	0	0	1	1	1	0
							1	0	0	0	1
							1	0	0	1	1
							1	0	1	0	0
							1	0	1	1	1
							1	1	0	0	1
							1	1	0	1	0
							1	1	1	0	0
							1	1	1	1	1

Demonstration of Logic Gates in Simple Circuits:

In this part, we are going to have some fun with simple circuits. By utilizing just 2 switches, a lamp and a battery, we explore the behaviors of a 2-input AND gate and a 2-input OR gate. These hands-on examples offer a concrete insight into how logic gates process signals and make logical decisions, forming the basis for more complex digital systems.

DO YOU KNOW?

The XNOR gate produces a high output (1) when an even number of its inputs are high (1).

I. AND Gate Demonstration:

In this demonstration, we will use following components as shown in the circuit diagram:

- 2 switches (Switch A and Switch B)
- 1 lamp
- 1 battery

The presented circuit exhibits two switches (A & B) and a bulb (F) arranged in series with a cell. According to basic circuit theory, the bulb illuminates only when a closed circuit permits the flow of current.

Behavior: The lamp will light up only when both Switch A and Switch B are in the ON position. If either or both switches are OFF, the lamp remains off.

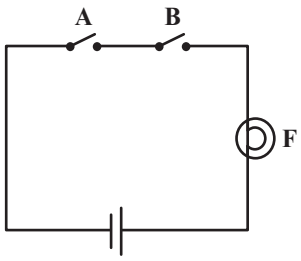


Figure 23. 9(a)
AND gate circuit demonstration

However, in the context of digital electronics, we approach the analysis differently. Instead of focusing on the physical flow of current, we interpret the circuit as making decisions, determining whether the output represents logic 1(the bulb is lit) or logic 0 (bulb not lit). This decision is contingent upon the logical states of the inputs (i.e., the switches), where an open switch corresponds to logic 0 and a closed switch corresponds to logic 1.

The truth table and the equivalent combination of above circuit in logic gate is shown in figure 23.9.



A	B	F
0	0	0
0	1	0
1	0	0
1	1	1

Figure 23. 9(b)
AND gate and its truth table

II. OR Gate Demonstration:

In this demonstration, we will use following components as shown in the circuit diagram:

- 2 switches (Switch A and Switch B)
- 1 lamp
- 1 battery

Figure 23.10 shows a simple circuit which contains two switches in parallel.

Behavior: The lamp will light up if either or both Switch A and Switch B are in the ON position. The lamp remains off only when both switches are OFF.

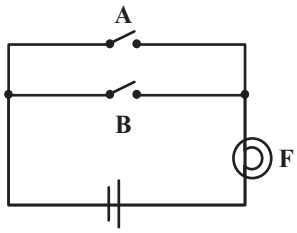


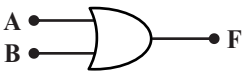
Figure 23. 10(a) OR gate circuit demonstration

The truth table and the equivalent combination of above circuit in logic gate is shown in fig 23.10

In these circuits:

- For the AND gate, the output is HIGH (lamp on) only when both inputs are HIGH.
- For the OR gate, the output is HIGH (lamp on) when at least one input is HIGH.

These simple circuits demonstrate the basic principles of AND and OR gates, showcasing how they process signals based on logical operations.

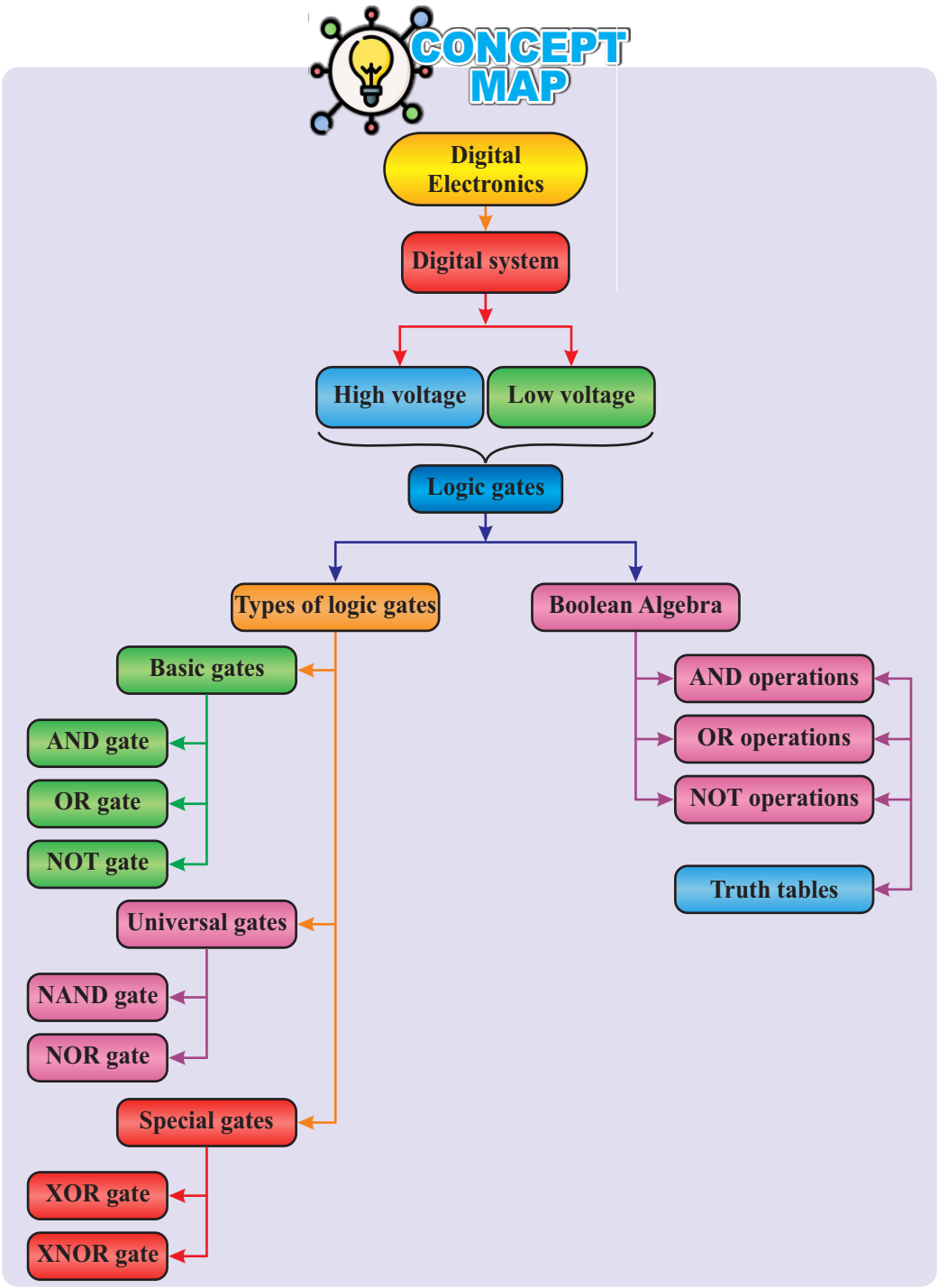


A	B	F
0	0	0
0	1	1
1	0	1
1	1	1

Figure 23.10(b) OR gate and its truth table

SUMMARY

- ✓ **Digital electronics** deals with digital signals represented by binary digits (0s and 1s).
- ✓ **Digital signals** use discrete voltage or current levels to represent information.
- ✓ **Common digital signal levels** include high level (logic 1), low level (logic 0), threshold level, and voltage swing.
- ✓ **Logic gates** are fundamental building blocks of digital circuits that perform logical operations on binary data.
- ✓ **Basic logic gates** include AND, OR, and NOT gates.
- ✓ **NAND and NOR** gates are universal gates that can be used to create any other logic gate.
- ✓ **XOR and XNOR** gates are special logic gates used for specific operations.
- ✓ **Logic gates can be implemented using transistors** which act as electronic switches.
- ✓ **Simple circuits with switches and bulbs** can be used to demonstrate the behavior of AND OR gates.


CONCEPT MAP
Digital Electronics
Digital system
High voltage
Low voltage
Logic gates
Types of logic gates
Boolean Algebra
Basic gates
AND gate
OR gate
NOT gate
Universal gates
NAND gate
NOR gate
Special gates
XOR gate
XNOR gate
AND operations
OR operations
NOT operations
Truth tables

EXERCISE

Section (A): Multiple Choice Questions (MCQs)

Choose the correct answer:

1. The voltage level, typically associated with a 'low' state in digital electronic is:
   (a) -5 volts (b) 0 volts (c) +5 volts (d) +10 volts
2. In binary representation, the correspond of '0' in digital electronics is:
   (a) Low state (b) High state (c) Open circuit (d) Closed circuit
3. The primary purpose of a truth table in digital electronics is to:
   (a) Determine input voltage
   (b) Measure circuit resistance
   (c) Analyze output states based on input combinations
   (d) Calculate circuit power
4. Which logic gate combination produces an output of '1' when at least one input is '1'?
   (a) AND (b) OR (c) XOR (d) NOT
5. In a circuit with a 2-input OR gate, the status of the lamp if either switch is closed
   (a) OFF (b) ON (c) Blinking (d) Flickering
6. The behavior of 2 input AND gate, when both switches are closed, Lamp is
   (a) OFF (b) ON (c) blinking (d) flickering
7. In a 3-input OR gate, if all inputs are '0'sothe output will be:
   (a) 1 (b) 0 (c) Undefined (d) Both A and B
8. If the input is '1', The output of a NOT gate will be:
   (a) 1 (b) 0 (c) Undefined (d) Both A and B
9. In a 2-input XOR gate, when both inputs are '1' than the output will be
   (a) 1 (b) 0 (c) Undefined (d) Both A and B
10. Which combination of logic gates is equivalent to an XOR gate?
    (a) AND + OR (b) NOT + NOR (c) OR + NAND (d) AND + NAND

Section (B): CRQs (Short Answered Questions):

1. Explain the significance of signal levels in digital electronics and how they are represented in terms of voltage.
2. Describe the binary representation used in digital electronics. How is this representation related to the 'high' and 'low' states?
3. Choose any three logic gate symbols and explain their operations.
4. Create a combination of logic gates that mimics the behavior of an XOR gate.

Section (C): ERQs (Long Answered Questions):

1. Design a simple circuit using a 2-input AND gate, a switch, a lamp, and a battery. Explain how the circuit operates and the conditions under which the lamp will be illuminated.
2. Construct a truth table for a 3-input OR gate. Explain how the truth table represents the relationship between input combinations and the resulting output.

Unit 24 Relativity

Teaching Periods 07 **Weightage %** 05



The image depicts a black hole's gravitational pull causing time dilation, where time appears to slow down near the event horizon. A clock near the black hole is shown with slower-moving hands, while a clock farther away remains normal, illustrating the relative nature of time. The black hole's warped spacetime is represented by a swirling, curved background, and a nearby planet or object is shown in a distorted, time-dilated state, highlighting the extreme gravitational effects on time and space.

In this unit student should be able to:

- Describe Relative Motion with suitable examples (same and opposite direction)
- Distinguish between inertial and non-inertial frames of reference
- Predict the motion of an object relative to a different frame of reference e.g. dropping a ball in a moving vehicle observed from the vehicle and by a person standing on the side walk.
- Analyze and evaluate the evidence confirming or denying Einstein's two postulates
- Identify that if c is constant then space and time become relative.
- Explain qualitatively and quantitatively the consequence of special relativity in relation to the relativity of simultaneity length contraction time dilation mass increase the equivalence between mass and energy
- Discuss the limitation on the maximum velocity of a particle imposed by special relativity
- Explain the implications of mass increase, time dilation and length contraction for space travel.
- Identify the role of special theory of relativity in global positioning, NAVSTAR system.
- Solve problems using $\Delta t = \frac{\Delta t_o}{\sqrt{1-\frac{v^2}{c^2}}}$, $L = L_o\sqrt{1-\frac{v^2}{c^2}}$, $m = \frac{m_o}{\sqrt{1-\frac{v^2}{c^2}}}$ and $E_o = mc^2$
- Describe the general relativity
- Understand gravity as space time continuum

Introduction:

Relativity, a fundamental theory in physics developed by Albert Einstein, revolutionized our understanding of space, time, and energy. It consists of two main parts: Special Relativity and General Relativity. Special Relativity, and the idea that time and space are relative, varying according to the observer's motion. General Relativity, introduced extends these ideas to include gravity, describing it not as a force but as the curvature of spacetime caused by mass and energy. These theories have profound implications for our understanding of the universe, influencing modern physics, cosmology, and even technologies like GPS.

DO YOU KNOW?

"Albert Einstein's theory of relativity (1905, 1916) transformed our understanding of space, time, and physics. This chapter covers the key concepts of Special Relativity and General Relativity, two groundbreaking theories that revolutionized our understanding of the universe."

Have you ever been in a car at a stoplight and felt like your car was rolling backward because the car next to you started moving forward? This sensation can be disorienting and is another everyday example of relative motion. How can we describe the motion of objects accurately when the observer's perspective changes?



By exploring relative motion, we can better understand and describe the motion of objects from various perspectives, whether they are in the same frame of reference or moving relative to one another. This chapter will delve into the principles of relative motion, providing a foundational understanding that is essential for studying physics. Through examples, case studies, and thought-provoking questions, we will uncover how motion is not absolute but relative to the observer's point of view.

24.1 Frame of References

A reference frame is a fundamental concept that helps us to describe and understand the relative motion of objects and events. It consists of a set of coordinate axes (usually x, y, and z axes) and a clock, allowing observers to precisely define where and when events occur.

A rigid framework (usually x, y and z-axes) relative to which the position and motion of an object can be measured is called frame of references.

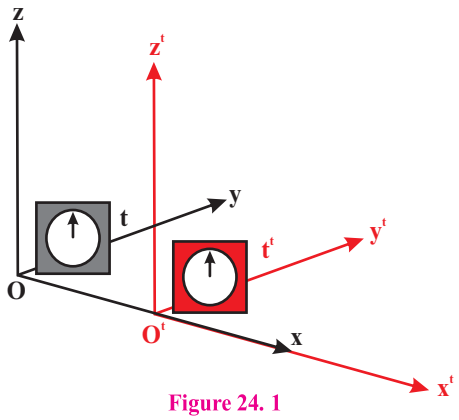


Figure 24. 1
Cartesian Co-ordinate System

The familiar Cartesian System of Co-ordinates, in which the position of the particle is specified by its three co-ordinates x, y, z, along the three perpendicular axes, is shown in Figure 24.1. We have indicated two observers

O and O′ and a particle P. These observers use frames of reference XYZ and X′Y′Z′, respectively. If O and O′ are at rest, they will observe the same motion of P. But if O and O′ are in relative motion, their observation of the motion of P will be different.

24.1.1 Relative Motion:

Relative motion is a concept that describes the motion of an object in relation to another observer or frame of reference. This viewpoint recognizes that motion is not absolute but depends on the observer's point of view.

Relative Motion: Same direction:

When two objects are moving in the same direction, their relative motion involves the difference in their velocities. The relative velocity is the velocity of one object as seen from the perspective of the other. The formula for relative velocity (v_{rel}) when the objects move in the same direction is given by:

$$v_{rel} = v_1 - v_2$$

where v_1 and v_2 are the velocities of the two objects, respectively.

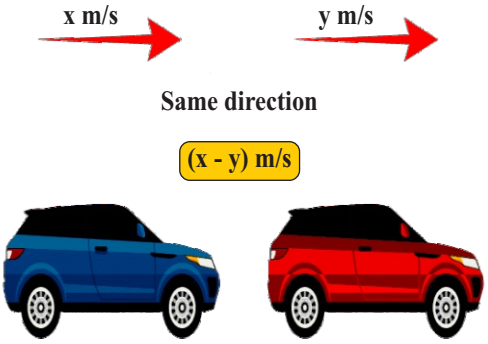


Figure 24.2 moving in same direction

Example:

Consider two cars, A and B, moving on a straight road. If car A has a velocity of 20 m/s, and car B has a velocity of 15 m/s in the same direction, the relative velocity of car B as observed from car A is

v_{rel}=20m/s – 15m/s=5m/s

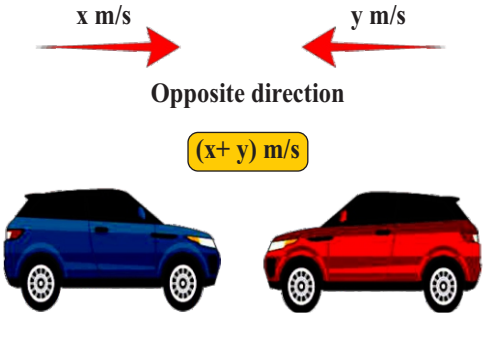


Figure 24.3 moving in different direction

Relative Motion: Opposite direction:

When two objects are moving in opposite directions, their relative motion involves the sum of their velocities. The formula for relative velocity (v_{rel}) when the objects move in opposite directions is given by:

$$v_{rel} = v_1 + v_2$$

Example:

Consider same two cars, A and B. But this time moving in opposite directions. Let's say car A is traveling at 80 m/s, and car B is now approaching it at 50 m/s.For a passenger on car A, car B appears to be moving towards them at relative velocity:

v_{rel}=80 m/s+50 m/s= 130 m/s

24.1.2 Types of Frames of References:

There are two types of frames of reference, Inertial frames of reference an dnon-inertial frames of reference.

Criteria	Inertial frame of Reference	Non-Inertial frame of Reference
Definition	A frame of reference with a constant	A frame of reference with an accelerating motion.
Motion of objects	Objects in uniform motion appear to follow straight lines or constant velocities.	Objects may appear to accelerate or experience fictitious forces.
Newton's first law	Newton's first law (Law of inertia) is valid in this frame.	Newton's first law is not valid due to accelerating motion.
Appearance of forces	Real forces are observed and can be directly measured.	Fictitious forces (e.g., centrifugal force, coriolis force) may appear due to the acceleration of the frame.
Equations of motion	Newton's laws of motion hold true in this frame.	Additional terms or transformations may be required to account of the frame.
Examples	A person inside a moving train, an object in free fall.	A person inside a spinning carousel, an object in circular motion.
Application	Often used in analyzing motion and dynamics in	Important in understanding phenomena in rotating systems, general relativity, and

Self-Assessment Questions:

1. Explain the difference between an inertial and a non-inertial frame of reference.
2. Provide an example of a commonly used frame of reference in everyday life.

24.1.3 Frames of references: Prediction of Motion:

Consider a vehicle moving at a constant speed v = 15 m/s with relative to the ground and a boy standing on the vehicle throws a ball with velocity v = 8 m/s in the same direction as the vehicle's motion. As shown in fig 24.4. If someone on the ground measures the ball's speed, it intuitively seems like it should be 23 m/s (15+8). Conversely, if you throw

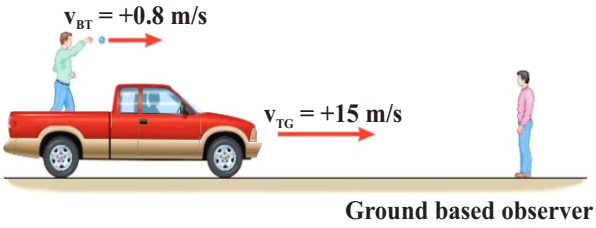


Figure 24.4: Moving and stationary frames of references

the ball in the opposite direction of the vehicle's speed, the ground observer would measure a speed of 7 m/s (15 − 8). Both observers are measuring the speed of the same object, but their relative motion leads to different results.

Now, let's derive a general equation applicable to any speeds. In our example, the vehicle's speed of 15 m/s becomes the relative speed of the two reference frames, denoted as u. The speed of the thrown ball is v′ (in the train's frame), and the measured speed on the ground is v. In a more formal way, we can express this relationship as:

$$v = v' + u$$

This equation is essentially the **Galilean transformation** of speeds. But, it is only applicable in mechanics.

Relative Motion: Galilean Transformation Equations:

The Galilean transformation equations describe the relationship between the coordinates of an event in two different inertial frames of references. These equations were formulated by Galileo and are applicable when the relative velocities involved are much smaller than the speed of light. With these equations, we can correctly measure the velocities of a ball in above example. Galileo concluded that:

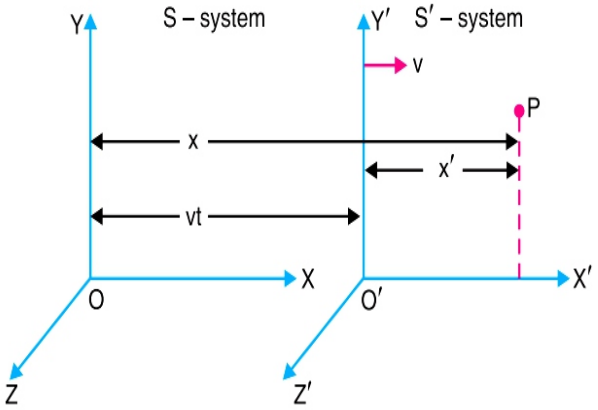


Figure 24. 5
Inertial frames of references: moving and stationary

"The laws of mechanics must be valid in all inertial frames of reference."

This is also known as the "Principle of Galilean Relativity".

Let x and x′ be two inertial frames (Figure 24.5). Let x be at rest and x′ moves with uniform velocity v along the positive X direction. We assume that v << c. Let the origins of the two frames coincide at t = 0. Suppose some event occurs at the point P. The observer O in frame x determines the position of the event by the coordinates x, y, z. The observer O′ in frame x′ determines the position of the event by the coordinate's x′, y′, z′. Let the time elapsed at the same rate in both frames, i.e., t = t′. There is no relative motion between x and x′ along the axes of Y and Z. Measurements in the X direction made is x frame will be greater than those made in x′ frame by the amount v_t, which is the distance x′ has moved in the X direction. Therefore,

$$x' = x - vt$$
$$y' = y$$
$$z' = z$$
$$t' = t$$

The set of above equations is known as the Galilean transformation equations.

24.2 Special Theory of Relativity:

Special theory of relativity explains how to interpret motion between different inertial frames of reference. In other words, it deals with the problems in which one frame of reference moves with a constant linear velocity relative to another frame of reference.

24.2.1 The Postulates of Special Relativity:

The special theory of relativity is based on two essential assumptions, commonly known as postulates.

Postulate I (Principle of Relativity):

The laws of Physics have the same form in all inertial frames of reference.

Consider a vehicle moving at a constant speed. Inside the vehicle, a passenger throws a ball straight up into the air. If we ignore any effects from the air, the passenger inside the moving vehicle sees the ball move up and then back down in a straight line, just as it would appear if someone standing still on the Earth threw a ball upwards. This means that the laws of physics that govern the motion of the ball, including gravity and equations for constant acceleration, work the same way whether the vehicle is moving or at rest.

Both observers, the one inside the vehicle and the one on the ground, see the ball go up and come back down. However, they see the path of the ball differently. The person on the ground sees the ball's path as a curved shape (a parabola), while the person in the vehicle sees it as a simple up-and-down motion. As shown in figure 24.6.

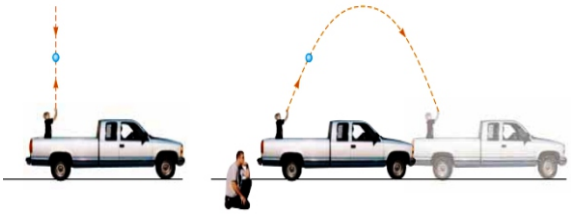


Figure 24.6: Principle of Relativity

Additionally, the person on the ground thinks that the ball has a horizontal component of velocity, which is the same as the vehicle's velocity. For the outside observer, the distance traveled along the parabolic path is longer than the path straight down but the time for the fall is the same. For the outside observer, the average velocity is greater.

Postulate II (Constancy of the speed of light):

The speed of light in vacuum has the same value, c=3 × 10^8m/s in all inertial reference frames, regardless of the velocity of the observer or the velocity of the source emitting the light.

Consider again the same example of throwing a ball in moving vehicle. As shown in figure 24.7, But this time, instead of throwing a ball, we shine a flashlight. If we were to do the same thing with light as we did with the ball, common sense might suggest that the

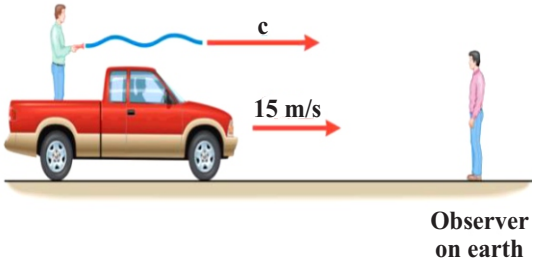


Figure 24.7 Two different frames of reference but speed of light is constant for both frames of references

speed of light would increase if the vehicle is moving in the same direction as the light according to Galilean Transformation Equation: v = c + u (if we replace ball's speed v′ with speed of light c). In fact, such an increase in the speed of light has never been found. In fact, in experiments carried out to test for the effect of the movement of the source on the speed of light (Michelson-Morley), the results indicate that the speed of light is completely unaffected by the motion of the source. It appears that the speed of light in a vacuum is constant regardless of relative motion.

In order to correct the Galilean equation of velocity addition, we have to consider velocity of light too. This correction was done by Hendrik Lorentz. The resultant equations are known as **Lorentz transformation equations**.

Relative Motion: Lorentz Transformation Equations:

We have to introduce new transformation equations which are consistent with the new concept of the invariance of light velocity in free space.

The Lorentz transformation equations for time and space coordinates (x,y,z,t) can be written as follows:

$$t' = \gamma\left(t - \frac{vx}{c^2}\right)$$
$$x' = \gamma(x - vt)$$
$$y' = y$$
$$z' = z$$

where t and x,y,z are the time and space coordinates in the original frame, t′ and x′,y′,z′ are the time and space coordinates in the frame moving with velocity v relative to the original frame as shown in figure 24.8, c is the speed of light, and $\gamma = \frac{1}{\sqrt{1-\frac{v^2}{c^2}}}$ is the Lorentz factor.

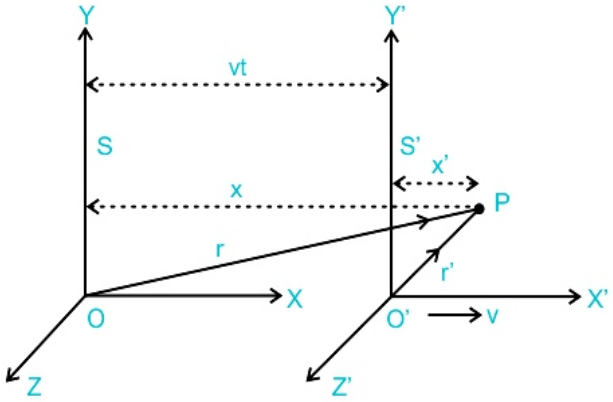


Figure 24.8 The Lorentz transformation

Self-Assessment Questions:

1. Calculate the value of Lorentz factor if the speed of car is 20m/s. Can we use Galilean Transformation equation for this value?
2. Why can't objects with mass exceed the speed of light?

24.2.2 Space and Time are Relative:

When the speed of light is constant, the time and space become relative concepts and are no longer absolute according to Newton.

The measurements of space and time intervals depend on the observer's relative velocity, leading to phenomena such as time dilation and length contraction. The constancy of the speed of light is a fundamental postulate in special relativity, challenging classical notions of absolute space and time.

24.2.3 Consequences of Special Theory of Relativity:

In Relativity, there is no such thing as an absolute length or absolute time interval. Here are some of the most important consequences of special relativity:

Relativity of Simultaneity:

The principle of relativity states that there is no preferred inertial frame of reference. It is defined as:

Two events that are simultaneous in one reference frame are in general not simultaneous in a second frame moving relative to the first. Simultaneity depends on the state of motion of the observer, and is therefore not an absolute concept.

Here's an example to illustrate this concept:

Imagine two observers, Boy A and B, standing on a train platform. They are equidistant from the center of the platform. At the exact moment when the train passes the center of the platform, lightning strikes both ends of the train. As shown in figure 24.9.

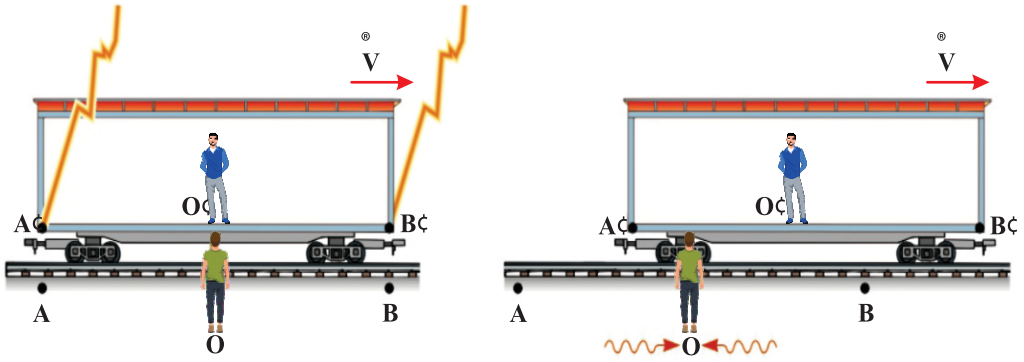


Figure 24. 9 Demonstration of Relativity of Simultaneity

From boy A's perspective:

Boy A sees the lightning strikes happen at the same time because she is stationary relative to the platform. Since light travels at a finite speed, the light from both lightning strikes reaches boy A simultaneously.

From boy B's perspective:

Boy B, however, is sitting on the moving train. Because the train is moving toward the lightning strike at the front of the train and away from the lightning strike at the back of the train, light takes longer to reach him from the back of the train than from the front.

As a result, Boy b perceives the lightning strike at the front of the train before the lightning strike at the back of the train. Thus, for boy B, the lightning strikes are not simultaneous.

This example demonstrates how the perception of simultaneity can differ between observers depending on their relative motion. In this case, what appears simultaneous to boy A (the lightning strikes) does not appear simultaneous to boy B due to his motion relative to the events.

Time Dilation:

The time interval, between two events occurring at a given point in the moving frame S′ appears to be longer to the observer in the stationary frame S. This effect is called time dilation.

Let Δt_o be the proper time measured by a clock that is at rest. The relative time t measured in another frame of reference is given by

$$\Delta t = \frac{\Delta t_o}{\sqrt{1-\frac{v^2}{c^2}}}$$

Where Δt is relativistic time and c is speed of light (3×10^8m/s). Here, the value of $\sqrt{1-\frac{v^2}{c^2}}$ should not be equal to zero; this will only occur when the speed of an object is less than thespeed of light c. If the speed of an object v equals c in the above equation, then Δt becomes infinite. This implies that time will seemingly stop, which is impossible. Therefore, no material object can travel at the speed of light. As shown in figure 24.10.

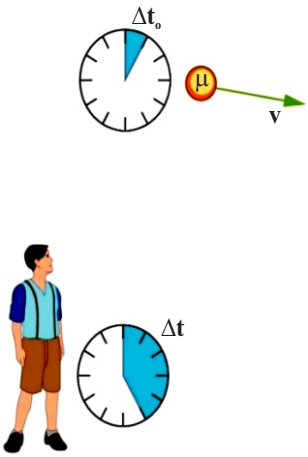


Figure 24. 10 Proper time elapsed by muon particle and relative time as measured by an observer

Length Contraction:

Let L_o be the length of a rod when the rod is stationary as shown in Figure 24.11 (a). If there is relative motion at speed v between an observer at rest and the rod along the length of the rod, then observer will calculate a relativistic length L given by

$$L = L_o\sqrt{1-\frac{v^2}{c^2}}$$

Where the length L_o is called proper length, L is relativistic length and c is speed of light (3×10^8m/s). The relative motion causes a length contraction as shown in Figure 24.11 (b). A greater speed v results in a greater contraction. Space contracts in only one direction, the direction of motion as shown in Figure 24.12.

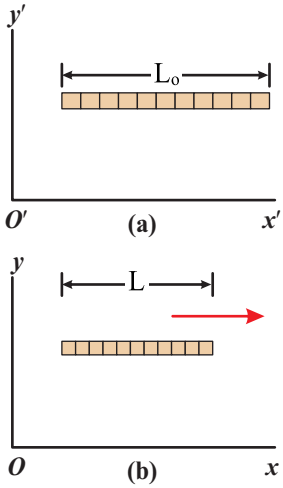


Figure 24. 11 Proper length and relative length of a rod moving with velocity v

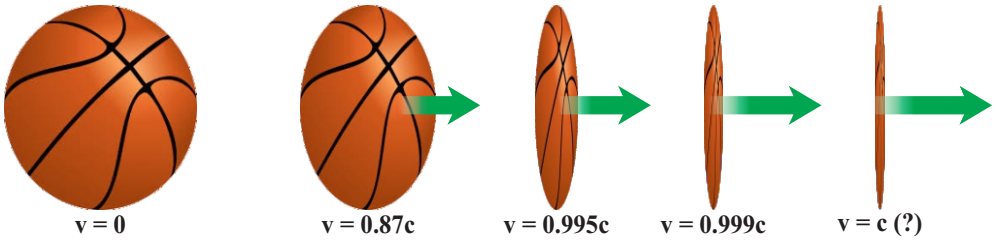


Figure 24. 12: Relative speed increases the contraction in only in one direction – the direction of motion. It doesn't affect other directions.

Mass Variation:

Let m_o be the rest mass of an object. If the object is moving at speed v then its relativistic mass m will be given by

$$m = \frac{m_o}{\sqrt{1 - \frac{v^2}{c^2}}}$$

Where m is relativistic mass and c is speed of light. The relative motion causes a mass variation.

Mass Energy Relationship:

According to Einstein's special theory of relativity, mass and energy are interchangeable. An object's mass m and the equivalent energy E_o are related by:

$$E_o = mc^2$$

where c is the speed of light. This equation gives the mass energy E_o that is associated with the object's mass m, regardless of whether the object is at rest or moving. If the object is moving, it has additional energy in the form of kinetic energy K. The total energy E is the sum of its mass energy and its kinetic energy:

$$E = E_o + K = mc^2 + K$$

$$E = \gamma mc^2$$

where γ is the Lorentz factor for the object's motion. And E is the total relativistic energy of the object. From above equations, relative kinetic energy can be calculated as:

$$K = E - E_o = E - mc^2 = \gamma mc^2 - mc^2$$

$$K = mc^2(\gamma - 1)$$

DO YOU KNOW?

Tachyons are hypothetical particles that travel faster than light. According to Einstein's theory of relativity, objects with real mass cannot reach the speed of light. Photons, which have no mass, can travel at this speed. For an object to travel at light speed, it would need imaginary mass. As a particle exceeds the speed of light, the denominator in the equation

$$m = \frac{m_o}{\sqrt{1 - \frac{v^2}{c^2}}}$$

Becomes imaginary. Imaginary mass would counteract this, making the particle appear to have real mass in its rest frame, while always moving faster than light.

24.2.4 Maximum Velocity of A Particle:

A basic result of the special theory of relativity is that the speed of an object cannot equal or exceed the speed of light. That the speed of light is a natural speed limit in the universe can be seen from above equations of mass variation, length contraction, time dilation and mass-energy relationship. If an object is accelerated to greater speeds, its mass becomes larger and larger. Indeed, if v were equals to c, the denominator in the equation would be zero and the mass would become infinite. To accelerate an object up to v=c would thus require infinite energy, and so it is not possible.

Time Dilation:

Special relativity also predicts time dilation, meaning that as an object approaches the speed of light, time appears to slow down for the object relative to an observer at rest. This effect becomes more pronounced as the object's velocity approaches c. Consequently, from the perspective of an observer, it would take an infinite amount of time for a massive object to reach the speed of light.

Length Contraction:

Another consequence of special relativity is length contraction, where objects moving at relativistic speeds appear shorter in the direction of motion when observed from a stationary frame. This effect prevents objects from achieving relativistic velocities within a finite distance because, as the object's velocity increases, its length contracts in the direction of motion, making it increasingly difficult to accelerate further.



Self-Assessment Questions:

1. If a rectangular box is moving with the speed of 0.9c. Will it shrink to smaller one in all directions? Discuss.

24.2.5 Relativistic Effects in Space Travel:

The implications of mass increase, time dilation, and length contraction for space travel are profound and have significant consequences for our ability to explore the universe at relativistic speeds. Here's how each of these effects impacts space travel:

Mass Increase:

- The implication for space travel is that as spacecraft approach relativistic speeds, their mass increases exponentially. This makes it increasingly difficult to accelerate them further, requiring enormous amounts of energy.
- Overcoming the mass increase becomes a significant engineering challenge for spacecraft propulsion systems. Current propulsion technologies, such as chemical rockets, would become impractical at relativistic speeds due to the massive energy requirements.

Time Dilation:

- For space travelers moving at relativistic speeds, time dilation means that their perception of time differs from that of stationary observers. A journey that may take several years according to Earth-based observers could be experienced as much shorter by the travelers due to time dilation.
- Time dilation has implications for interstellar travel, where long-duration missions could be undertaken without experiencing the full effects of time passing, as perceived by Earth-based observers. However, it also presents challenges for communication and synchronization with mission control on Earth.

Length Contraction:

- For space travelers, length contraction means that distances along their direction of motion appear shorter. This has implications for navigation and spacecraft design, as distances may be perceived differently by travelers moving at relativistic speeds compared to stationary observers.
- Additionally, length contraction affects the perception of space travel distances. What may appear as a vast distance to stationary observers could be contracted from the perspective of travelers moving at relativistic speeds.

24.2.6 Navigation Satellite Timing and Ranging (NAVSTAR) SYSTEM):

It is the official name for the Global Positioning System (GPS). It's a network of 24 satellites orbiting Earth that allows us to determine our location on the planet with incredible accuracy. As shown in figure 24.13 (a,b).

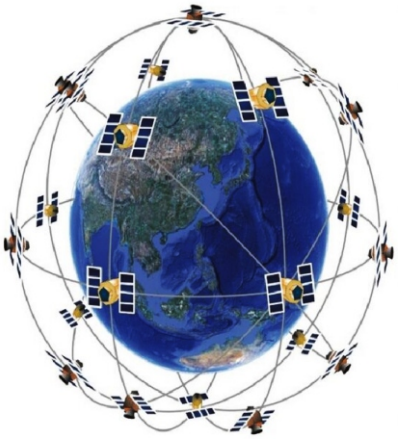


Figure 24.13 (a) NAVSTAR SYSTEM

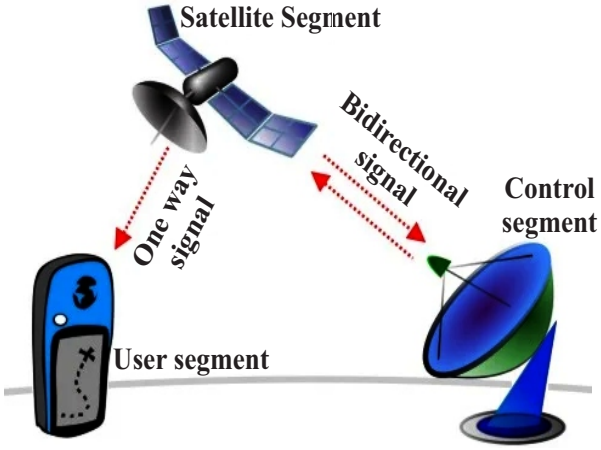


Figure 24. 13 (b): The NAVSTAR Global Positioning System (GPS)

One of the core principles of NAVSTAR system is synchronized clocks on the satellites. These clocks tick incredibly accurately, but due to the effects of relativity, this isn't as simple as it seems. Here's why:

Time Dilation:

As the satellites zoom around Earth at high speeds (roughly 11,000 km/h), their clocks actually run slightly slower compared to clocks on Earth.

Gravity:

The satellites being closer to Earth experience a stronger gravitational pull, which, according to General Relativity, also slows down their clocks compared to those on the ground.

If these effects weren't accounted for, our GPS calculations would be significantly off. But, because scientists understand space-time and its relativistic effects, they can account for this time discrepancy and maintain the accuracy of GPS positioning.

Using Equation of time dilation and the speed of the GPS satellites, we can calculate the difference between the dilated time interval and the proper time interval as a fraction of the proper time interval and compare the result to the stability of the GPS clocks:

$$\frac{\Delta t - \Delta t_o}{\Delta t_o} = \frac{1}{\sqrt{1 - v^2/c^2}} - 1 = \frac{1}{\sqrt{1 - (4000)^2/(3 \times 10^8)^2}} - 1$$

$$= \frac{1}{1.1 \times 10^{10}}$$

This indicates the level of stability. It means the clock deviates by one part in 10 trillion (10^{13}) over a certain period. In other words, the clock's timekeeping is extremely precise, with an error margin of just 0.0000000001%.

Worked Example 24.1

The period of a pendulum is measured to be 3.00 s in the inertial frame of the pendulum. What is the period as measured by an observer moving at a speed of 0.950c with respect to the pendulum?

Solution:

Step 1: Write down the formula and rearrange if necessary.

Using the equation of mass variation:

$$t = \frac{t_o}{\sqrt{1 - \frac{v^2}{c^2}}}$$

Step 2: Substitute the proper time and relative speed in the equation of time dilation:

$$t = \frac{t_o}{\sqrt{1 - \frac{v^2}{c^2}}} = \frac{3}{\sqrt{1 - \frac{0.95c^2}{c^2}}} = 9.61s$$

Result: The moving observer considers the pendulum to be moving, and moving clocks are observed to run more slowly: while the pendulum oscillates once in 3 s for an observer in the rest frame of the clock, it takes nearly 10 s to oscillate once according to the moving observer.

Worked Example 24.2

A starship is measured to be 125 m long while it is at rest with respect to an observer. If this starship now flies past the observer at a speed of 0.99c, what length will the observer measure for the starship?

Solution:

Step 1: Write down the formula and rearrange if necessary.

Using the equation of mass variation:

$$L = L_o\sqrt{1 - \frac{v^2}{c^2}}$$

Step 2: Substitute the proper length and relative speed in the equation of length contraction:

$$L = L_o\sqrt{1 - \frac{v^2}{c^2}} = 125 \times \sqrt{1 - \frac{0.99c^2}{c^2}} = 17.6m$$

Result: $L = 17.6m$

Worked Example 24.3

At what speed will an object's relativistic mass be twice its rest mass?

Solution:

Step 1: Write down the formula and rearrange if necessary.

Using the equation of mass variation: $m = \frac{m_o}{\sqrt{1-\frac{v^2}{c^2}}}$

Step 2: Where its relativistic mass m = $2m_o$. Then velocity of the object will be:

$$2m_o = \frac{m_o}{\sqrt{1-\frac{v^2}{c^2}}} \quad = 2 \times \sqrt{1-\frac{v^2}{c^2}} = 1$$

Step 3: Squaring both sides:

$$1-\frac{v^2}{c^2} = \frac{1}{4}$$

$$1-\frac{1}{4} = \frac{v^2}{c^2}$$

$$v = \frac{\sqrt{3}c}{2}$$

Result: This speed is approximately 86.6% of the speed of light.

24.3 General Theory of Relativity:

Einstein worked on the fact that acceleration produced the same effect as gravitation. For example, if you were in an accelerating spaceship, or just an elevator, you could not tell if the force on you was from inertia or gravitation. While Newton's Law of Universal Gravitation works well for ordinary gravitational fields, it is inaccurate when the gravitational intensity is high.

27.3.1 General Theory of Relativity:

Albert Einstein's general theory of relativity is a new way to explain gravity. According to Newton, tossed balls curve because of a force of gravity. According to Einstein, tossed balls and light don't curve because of any force, but because the space time in which they travel is curved. General relativity describes how gravity is caused by the curvature of space-time.

The general theory of relativity is based on the following two postulates.

Spacetime is a four-dimensional continuum that includes three dimensions of space and one dimension of time. Imagine it as a two-dimensional rubber sheet. If you place a heavy object, like a bowling ball, on the sheet, it will cause the sheet to deform or bend around the object. Now, if you roll a smaller ball across the sheet, it will move along the curved paths created by the heavy object. This is analogous to how planets and other celestial bodies move along the curved space-time around massive objects like the Sun. As illustrated in Figure 24. 14.

Postulate I:

The laws of physics may be expressed in equations having the same form in all frames of reference, regardless of their states of motion.

Thus, the general theory covers uniform as well as accelerated motion. Hence it is able to describe gravitational phenomena.

Postulate II (Principle of equivalence):

There is no way to spot the difference between gravity and acceleration for an observer in a closed laboratory

OR

There is no way for an observer in a closed laboratory to distinguish between the effects produced by a gravitational field and those produced by an acceleration of the laboratory, This postulate follows from the experimental observation that the inertial mass of a body is always exactly equal to its gravitational mass.

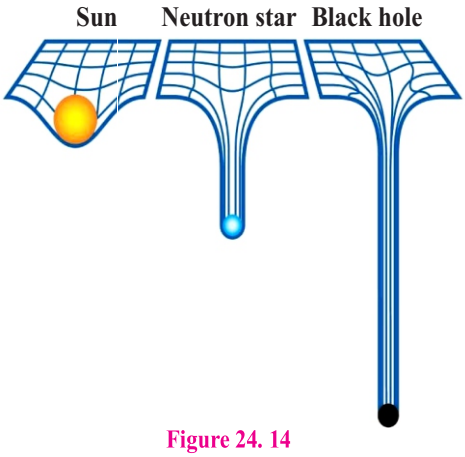


Figure 24. 14
Spacetime wrapping near heavy objects

General relativity has a number of consequences:

- Einstein's general theory of relativity states that time is a fourth dimension adding to thethree dimensions of space. Einstein called this four-dimensional geometry as space-time.
- The general theory of relativity also predicted light coming from a strong gravitational field would have its wavelength shifted toward longer wavelengths, called a red-shift.
- The theory also predicted that when gravity becomes great enough, it would produce objects called **black holes**. Black holes are objects whose gravity is so massive that light cannot escape from the surface at all. Since no light can escape, such objects would appear black
- Light is bent as it passes through curved space-time. This can cause distant objects to appear distorted or magnified. **Gravitational lensing** has been used to study distant galaxies and other objects.
- **The expansion of the universe:** General relativity predicts that the universe is expanding. This has been confirmed by astronomical observations.

DO YOU KNOW?

Dr. Katie Bouman captures the first image of a black hole. She and her team of NASA scientists developed a new technology to make the discovery possible. In 2019, astronomers observed the first direct evidence of existence of a black hole. (With first ever picture of black hole).



27.3.2 Gravity as Space Time Continuum:

According to General Theory of Relativity, when you put mass in space-time, it bends the geometry of space-time. According to this theory, gravity is not just a force between masses, as described by Newtonian physics, but rather a manifestation of c the curvature of space-time caused by mass and energy.

Here's a breakdown of what this concept entails:

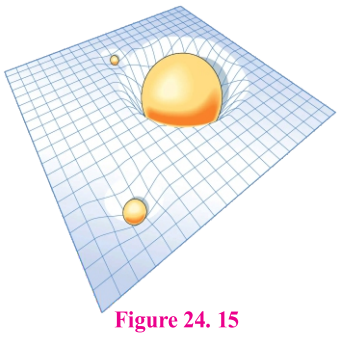


Figure 24. 15
Heavier object bends the geometry of spacetime more than lighter object

Space-time Continuum:

- Space and time are unified into a single, four-dimensional entity known as space-time. In the absence of gravity or significant mass, space time is flat.
- However, the presence of mass and energy warps or curves the fabric of space-time. This curvature is what we perceive as the force of gravity.

DO YOU KNOW?

Masses placed in space cause the space to be curved. Curved space causes masses moving in a straight line to follow a curved path. Curved space also causes time to run more slowly.

Curvature of Space-time:

- Massive objects, such as stars, planets, and galaxies, curve the space-time around them. The greater the mass or energy density, the greater the curvature of space-time.
- Objects move along paths dictated by the curvature of space-time, which we perceive as being influenced by gravity. For example, Earth orbits the Sun because the Sun's mass curves the space-time around it, causing Earth to follow a curved path.

Effects of Gravity:

- In the presence of gravity, objects follow the shortest possible paths, known as geodesics, through the curved space-time. These paths are not necessarily straight.
- Lines in the traditional sense but are instead curved trajectories determined by the gravitational field.
- The strength of gravity is determined by the curvature of space-time. The steeper the curvature, the stronger the gravitational force experienced by nearby objects.

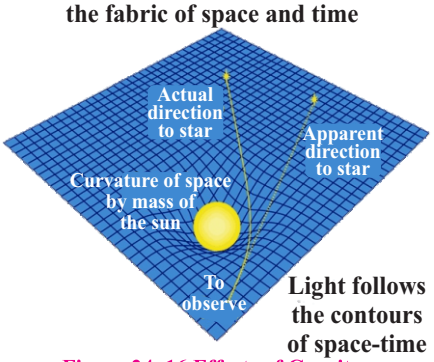


Figure 24. 16 Effects of Gravity

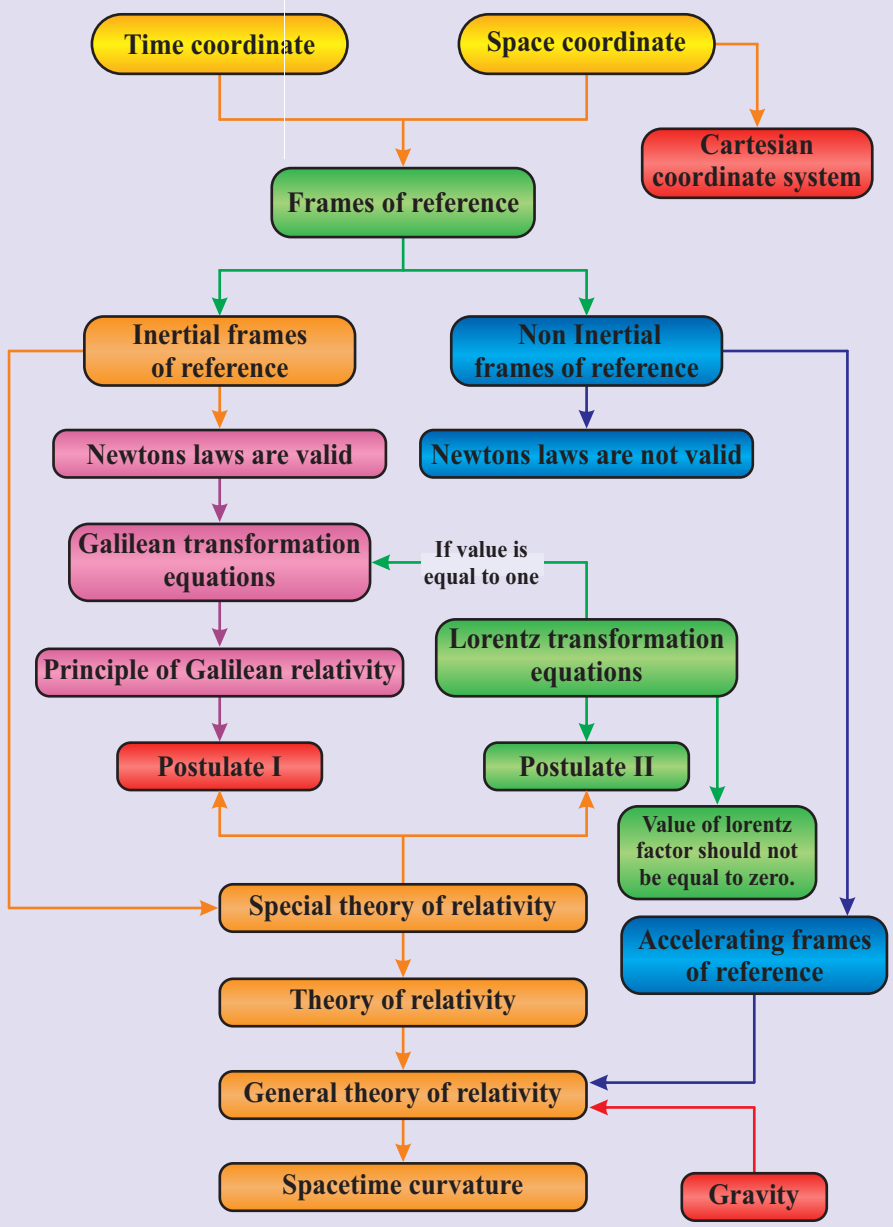
Physical Interpretation:

- Gravity is thus interpreted as the result of objects moving through the curved space-time created by mass and energy. Massive objects "warp" the space-time around them, causing other objects to move in response to this curvature.
- This concept provides a unified explanation for both the gravitational force and the motion of objects in space, incorporating both space and time into a single framework.



- ✓ Special Reactivity applies to objects moving at constant velocity relative to each other. It introduces the concept of reference frames and explains how the laws of physics are the same in all inertial reference frames.
- ✓ Two important postulates of Special Relativity are:
  - ➢ The laws of physics are the same in all inertial frames of reference.
  - ➢ The speed of light in a vacuum is the same for all observers regardless of their relative motion or the motion of the source of the light.
- ✓ Consequences of Special Relativity include:
  - ➢ Relativity of simultaneity: Two events that are simultaneous in one frame of reference may not be simultaneous in another frame moving relative to the first.
  - ➢ Time dilation: Time intervals between events appear longer for objects in motion relative to an observer at rest.
  - ➢ Mass variation refers to the change in the mass of an object as its speed approaches the speed of light. According to relativity, the mass of an object increases with its velocity, becoming infinite as it reaches the speed of light.
  - ➢ Length contraction: Objects in motion appear shorter in the direction of their motion relative to an observer at rest.
  - ➢ Mass-energy equivalence: Mass and energy are interchangeable $E=mc^2$.
- ✓ General Relativity applies to objects under any motion, including acceleration. It explains gravity as a curvature of space-time caused by mass and energy.
- ✓ Two postulates of General Relativity are:
  - ➢ The laws of physics may be expressed in equations having the same form in all frames of reference, regardless of their states of motion.
  - ➢ There is no way for an observer in a closed laboratory to distinguish between the effects produced by a gravitational field and those produced by an acceleration of the laboratory itself.
- ✓ Consequences of General Relativity include:
  - ➢ Space-time: Space and time are unified into a single four-dimensional fabric.
  - ➢ Gravitational red-shift: Light coming from a strong gravitational field has a longer wavelength (red-shifted).
  - ➢ Black holes: Extremely massive objects whose gravity is so strong that not even light can escape.
  - ➢ Gravitational lensing: Light bends as it travels through curved spacetime, which can distort, diminish or magnify distant objects.
  - ➢ Expansion of the universe: General relativity predicts the universe is expanding, which has been confirmed by astronomical observations.

CONCEPT MAP


Time coordinate
Space coordinate
Cartesian coordinate system
Frames of reference
Inertial frames of reference
Non Inertial frames of reference
Newtons laws are valid
Newtons laws are not valid
Galilean transformation equations
If value is equal to one
Lorentz transformation equations
Principle of Galilean relativity
Postulate I
Postulate II
Value of lorentz factor should not be equal to zero.
Special theory of relativity
Accelerating frames of reference
Theory of relativity
General theory of relativity
Spacetime curvature
Gravity



Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. The General Theory of Relativity was a new way of understanding of
   (a) The speed of light (b) gravity
   (c) mass (d) force
2. An object at rest has a mass of 1 kg. What is its mass when it is moving at a speed of 0.9c?
   (a) Infinite (b) 1.2 kg (c) 2.3 kg (d) 1 kg
3. The equivalence of principle in general relativity is:
   (a) The equivalence of inertial and gravitational mass
   (b) The equivalence of electric and magnetic fields
   (c) The equivalence of space and time
   (d) The equivalence of matter and energy
4. Which of the following phenomena is NOT predicted by special relativity?
   (a) Time dilation (b) Length contraction
   (c) Gravitational waves (d) Relativistic mass increase
5. What does the equivalence principle state about acceleration and gravity?
   (a) They are completely different forces (b) They are indistinguishable for an observer
   (c) Gravity is stronger than acceleration (d) Acceleration cancels out gravity
6. Light passing near a massive object like a star will bend due to:
   (a) Gravitational lensing (b) refraction
   (c) reflection (d) diffraction
7. What is the main reason astronomers cannot directly observe black holes?
   (a) They are too small (b) They deflect light
   (c) They are too far away (d) their immense gravity traps light
8. Objects cannot exceed the speed of light because:
   (a) Their mass becomes infinite (b) Their length becomes zero
   (c) Time slows down to zero (d) they lose all energy
9. Why is the Galilean transformation not valid at high speeds?
   (a) It cannot handle accelerating frames
   (b) It violates the constancy of the speed of light
   (c) It only works in flat spacetime
   (d) It neglects time dilation effects
10. The example of inertial frame in our everyday life is:
    (a) A car accelerating on a highway
    (b) A train moving smoothly at constant speed
    (c) A person standing on a spinning platform
    (d) an airplane encountering turbulence

Section (B): CRQs (Short Answered Questions):

1. Show that for values of v << c, Lorentz transformation reduces to the Galilean transformation.
2. If a particle could move with the velocity of light, how much K.E. would it possess?
3. Explain the difference between Special and General Relativity in simple terms.
4. Differentiate between Inertial Frames of Reference and Non- Inertial Frames of Reference.
5. Why can't any object move at the speed of light?
6. What is the limitation in the Galilean Transformation Equation, and how did Lorentz solve it?
7. Calculate the value of γ (Lorentz factor) if the object is moving at the speed of light.

Section (C): ERQs (Long Answered Questions):

1. State and explain the basic postulates of Einstein's special theory of relativity. Discuss length-contraction, mass variation and time-dilation.
2. Explain the concept of mass-energy equivalence. Derive Einstein's mass-energy relation and demonstrate that 1 atomic mass unit (u) is equivalent to 931 MeV.
3. Discuss the important conclusions derived from General Theory of relativity. What are the experimental observations in favour of these conclusions?
4. How does it principle of relativity differ from the classical Galilean view?
5. Explain the concept of spacetime curvature in general relativity and how it is used to visualize the effects of gravity.
6. Derive the basic equations of the Galilean transformation and explain how they relate the positions and velocities of objects in different inertial frames.
7. Discuss the Lorentz transformation equations in special relativity and how they describe the relationship between space and time coordinates in different inertial frames. Give examples to illustrate their application.

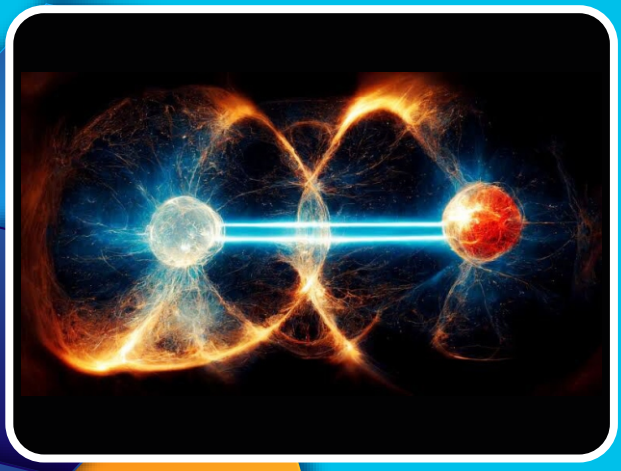
Section (D): Numerical:

1. A rod 1 meter long is moving along its length with a velocity 0.6c. Calculate its length as it appears to an observer (a) on the earth (b) moving with the rod itself.
   [Ans: (a) 0.8m, (b) 1m]
2. How fast would a rocket have to go relative to an observer for its length to be contracted to 99% of its length at rest? **[Ans: $42.45 \times 10^6 m/s$]**
3. A particle with a proper lifetime of 1μs moves through the laboratory at 2.7×10^8 m/s. (a) What is its lifetime, as measured by observers in the laboratory? (b) What will be the distance traversed by it before disintegrating? **[Ans: (a) $2.3 \times 10^{-6} s$, (b) 620m]**
4. At what speed is a particle moving if the mass is equal to three times its rest mass?
   [Ans: $\frac{2\sqrt{2}}{3}c$]

5. If 4 kg of a substance is fully converted into energy, how much energy is produced? **[Ans: 3.6×10^{17}J]**
6. Calculate the rest energy of an electron in joules and in electron volts. **[Ans: 8.2×10^{-14}J, 0.511MeV]**
7. Calculate the K.E. of an electron moving with a velocity of 0.98 times the velocity of light in the laboratory system. **[Ans: 3.3×10^{-13}J]**
8. At what velocity does the K.E. of a particle equal its rest energy? **[Ans: $\frac{\sqrt{3}}{2}c$]**
9. A particle of rest mass m_0 moves with speed $c/\sqrt{2}$. Calculate its mass, momentum, total energy and kinetic energy. **[Ans: $\sqrt{2}m_0$, m_0c, $\sqrt{2}\ m_0c^2$, $0.41\ m_0c^2$]**
10. The nearest star to Earth is Proximal Centauri, 4.3light- years away. A spaceship with a constant speed of 0.800c relative to the Earth travels toward the star.
    (a) How much time would elapse on a clock as measured by travelers on the spacecraft? **[Ans: (a) 3.22 years]**
    (b) How long does the trip take according to Earth observers? **[Ans: (b) 5.38 years]**

Unit 25 Quantum Physics

Teaching Periods 13 **Weightage %** 09



Quantum entanglement is a phenomenon in which two or more particles become correlated in such a way that the state of one particle cannot be described independently of the others, even when they are separated by large distances. Two particles connected by a wavy line, with arrows indicating their correlation Quantum entanglement has fascinating implications for quantum computing, cryptography, and our understanding of reality itself!

In this unit student should be able to:

- Describe the concept of black body radiation.
- Describe how energy is distributed over the wavelength range for different temperatures.
- Describe Planck's hypothesis that radiation emitted and absorbed by the walls of a black body cavity is quantized.
- Elaborate the particle nature of electromagnetic radiation.
- Solve problems by using $E = hc/\lambda$
- Describe the phenomenon of photoelectric effect.
- Explain photoelectric Effect on the basis of quantum theory.
- Solve the problems by using$hf - \varphi = KE$.
- Identify data sources, gather, process and present information to summarize the use of the photoelectric effect in solar cells and photocells.
- Explain the particle model of light in terms of photons with particular energy and frequency.
- Describe Compton's effect qualitatively.
- Solve problems by using $\Delta\lambda = h/moc\ (1 - Cos\theta)$
- Explain the process of pair production on the basis of conservation laws.
- Describe conservation laws in the annihilation of matter.
- Describe the impact of de Broglie proposal that any kind of particle has both wave and particle properties.
- Describe the confirmation of de Broglie proposal by Davisson and Germer experiment in which the diffraction of electrons by the surface layers of a crystal lattice was observed.
- Explain how the very short wavelength of electrons, and the ability to use electric and magnetic fields to focus them, allows electron microscope to achieve very high resolution.
- Search and describe the role of electron microscope to study the micro structures and properties of matter.
- Describe uncertainty principle.

Introduction:

Quantum physics is the study of matter and energy at atomic and sub-atomic level. It aims to discover the properties and behaviors of the very building blocks of nature. At these scales, the classical laws of physics no longer apply, and strange, seemingly random phenomena govern the behavior of particles. In early 20th century, the field of quantum physics was developed as a result of certain experimental discoveries which could not be described by classical physics. The fundamental discovery of quantum physics was that energy is radiated as discrete packets, or quanta. Quantum physics explores the principles of wave-particle duality, uncertainty, superposition and quantization, which form the basis of our understanding of the behavior of matter, energy, and the fabric of reality itself.

25.1 Quantum Theory of Radiation:

It was introduced by Max Planck and Albert Einstein It explains how energy is emitted and absorbed in discrete packets (quanta), the key points of Quantum theory of radiation are:

- Energy is quantized (small packets, not continuous)
- Frequency (not amplitude) determines energy of radiation
- Radiation has both wave and particle properties (duality)

This theory revolutionized our understanding of light and radiation!

25.1.1 Black Body Radiation:

A black body is an idealized object which absorbs all incident radiation and also emits it in a continuous spectrum of colors, depending upon its temperature as illustrated in fig 25.1. This evidence encouraged physicists to explain the mechanisms through which entities like stars and heated metals produce both light and heat.

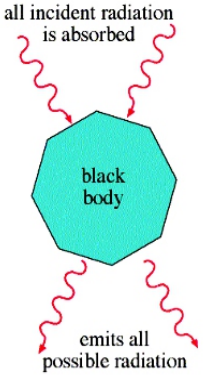
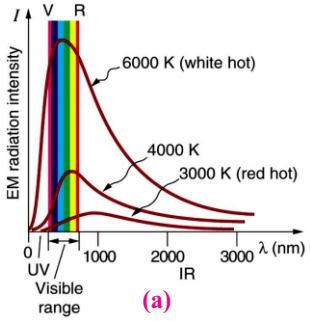


Figure 25.1
Black body Radiation



(a)



(b)

Figure 25.2 (a, b) Radiation intensity and red-hot horse

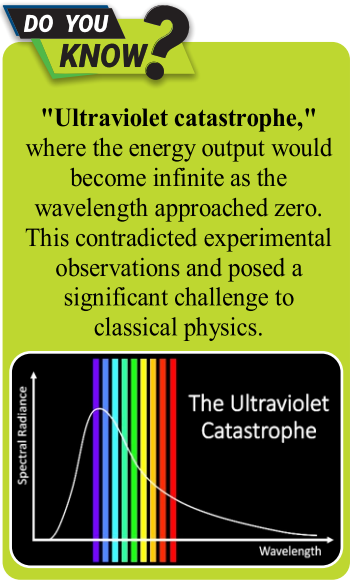
The spectrum of light emitted by heated objects was unexplained until 19th century. At normal temperature, we are not aware of this electromagnetic radiation due to low intensity. At higher temperatures, infrared radiation is emitted that we can feel heat if we are close to the object. At higher temperatures, objects actually glow, such as red-hot horse-shoe as shown in figure 25-2 (b). At temperatures more than 2000 K, objects glow with a yellow or whitish

color, such as white-hot iron. The light emitted contains a continuous range of wavelengths or frequencies, and the spectrum is a plot of intensity versus wavelength or frequency. As the temperature increases, the electromagnetic radiation emitted by objects not only increases in total energy but has its peak intensity and higher frequencies. Figure 25.2 Graphs of blackbody radiation (from an ideal radiator) at three different radiating temperatures.

The intensity or rate of radiation emission increases directly with temperature, and the peak of the spectrum shifts toward the visible and ultraviolet parts of the EM spectrum. The shape of the spectrum cannot be described with classical physics.

Some experimental facts about blackbody radiations:

- The blackbody spectrum depends only on the temperature of the object, and not on what it is made of i.e., material. An iron horseshoe, a ceramic vase, and a piece of charcoal --- all emit the same blackbody spectrum if their temperatures are the same. As the temperature of an object increases, it emits more intense blackbody radiations of all wavelengths.
- As the temperature of an object increases, the peak wavelength of the blackbody spectrum curve shifts towards shorter wavelength (Blue-shift). For example, blue stars are hotter than red stars.
- The blackbody spectrum peak shifted always becomes small at the left-hand side i.e., the short wavelength, high frequency.



25.1.2 Classical explanation of black body radiation:

Rayleigh-Jeans Law:

This Law states that the energy per unit volume per unit wavelength of blackbody radiation is inversely proportional to the fourth power of the wavelength (λ), not directly proportional to the square of the wavelength.

The law is given by the formula:

$$E(\lambda, T) = \frac{2ckT}{\lambda^4}$$

Mathematically expressed as:

$$E(\lambda, T) = \frac{Constant}{\lambda^4} \quad \ldots\ldots 25.1$$

where:

- $E(\lambda,T)$ is the spectral radiance
- λ is the wavelength,

The failure of the Rayleigh-Jeans Law highlighted the need for a new theoretical framework, eventually leading to Max Planck's development of the quantum theory of radiation in 1900. Planck's work successfully resolved the ultraviolet catastrophe and laid the foundation for quantum physics.

- The graph in figure depicts that for large

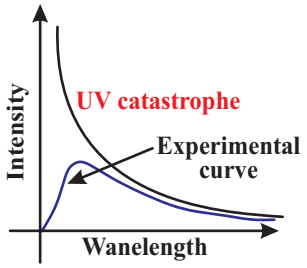


Figure 25.3 ultraviolet catastrophe

wavelengths it fitted the experimental data but it had major problems at shorter wavelengths.

Wien's Displacement Law:

This law breaks down completely at low frequencies. Wien's displacement law states that the wavelength at which the intensity of blackbody radiation is maximum is inversely proportional to the temperature of the blackbody. As the temperature increases, the peak of the blackbody radiation curve shifts to shorter wavelengths. This relationship is mathematically expressed as:

Λmax T=Constant

$$\lambda_{max} = \frac{Constant}{T} \ldots\ldots 25.2$$

Where λ_{max} is the wavelength at which a blackbody radiates at a given temperature T.

Stefan's Law:

The experimental relation is known as Stefan Boltzmann's law; it states that the total energy radiated per unit surface area of a black body is proportional to the fourth power of its temperature.

Mathematically, it is expressed as:

$P = \sigma AT^4$…….. 25.3

Where A is the surface area of a blackbody, T is its temperature and σ is the Stefan's-Boltzmann constant,

$\sigma = 5.67 \times 10^{-8}$ W / (m^2. K^4).



Stefan's law

It enables us to estimate how much energy a star is radiating by remotely measuring temperature.

25.1.3 Planck's Hypothesis:

According to Planck's Hypothesis, the energy levels of the oscillators in the walls of a blackbody are quantized, meaning they can only have specific, discrete values. This quantization of energy levels is what leads to the characteristic spectrum of blackbody radiation, which Planck's Law describes accurately.

Planck proposed that electromagnetic energy could only be emitted or absorbed in discrete packets, which he called "quanta." This assumption led to the concept of energy levels being quantized, meaning that the energy could only take on certain discrete values. The energy (E_n) of an oscillator in the cavity walls is given by the equation:

$$E_n = nhf \ldots\ldots\ldots\ldots 25.4$$

where: E_n is the energy of the oscillator is an integer (1, 2, 3, …), h is Planck's constant, f is the frequency of the oscillator.

This relationship, now known as Planck's Law, marked the birth of quantum theory, fundamentally changing our understanding of energy and leading to the development of quantum mechanics

An oscillator in the wall can receive energy from the radiation in the cavity (absorption), or it can emit energy to the radiation in the cavity (emission). The absorption process sends the oscillator to a higher quantum state, and the emission process sends the oscillator to a lower quantum state. Whichever way this exchange of energy goes, the smallest amount of energy allowed to exchange is $h\nu$. There is no upper limit to how much energy can

be exchanged, but whatever is exchanged must be an integer multiple of ***hν***. If the energy packet does not have this exact amount, it is neither absorbed nor emitted at the wall of the blackbody.

Self-Assessment Questions:

1. How does the energy distribution of black body radiation change with temperature? Describe the relationship between temperature and the peak wavelength of radiation.
2. Describe the ultraviolet catastrophe problem in the context of black body radiation. How did Max Planck's quantum hypothesis resolve this issue?

Worked Example 25.1

At the Earth's surface, the energy flux in sunlight is 1.02×10^3 watt/m^2. If a black sheet of paper faces the Sun, what is the equilibrium temperature of the paper? Assume that the bottom of the paper is insulated, so the only heat loss is by blackbody radiation from the top surface.

Solution:

Step 1: Energy = P / A = 1.02×10^3 watt/m^2

T =?

$\sigma = 5.67\times10^{-8}$ watt/m^2. K^4

Step 2: Write down the formula and rearrange if necessary.

$E = \sigma T^4$

Step 2: Put the values in above formula $\sigma T^4 = 1.02\times10^3$ $T = \left[\frac{1.02\times10^3}{5.67\times10^{-8}}\right]^{1/4}$

T = 362 K or

T = 90 °C

Result: T = 90 °C

25.1.4 Particle Nature of Electromagnetic Radiation:

Electromagnetic radiation can be described in terms of a stream of **mass-less particles, called photons**. Each photon contains a certain amount of energy. The different types of radiation are defined by the amount of energy found in the photons.

The particle nature of light states that light consists of particles called 'Photons' (Particles do not interfere) That is, when the space is occupied by some particle, other particles cannot occupy the same space. Experiments from the last century, such as the **photoelectric effect** and **atomic line spectra**, can only be explained if EM radiation is assumed to behave as particles. The energy of a photon depends upon the frequency as described in Planck's hypothesis.

$$E = h\nu$$

$$But\ f = \frac{1}{t}\ and\ V = \lambda\nu$$

$$c = \lambda\nu$$

$$E = \frac{h\nu}{\lambda} \ldots\ldots\ldots 25.5$$

h = Planck's constant = h=6.626×10^{-34}

c = Speed of light = 3×10^{8} m/s

λ = Wavelength of Photon particle

In EM spectrum as you move right wards the wavelength decreases and frequency increases; consequently, energy increases.

Worked Example 25.2

Calculate the energy of a photon with a frequency of 3×10^{18} Hz. Analyze the question
Determine the energy of a photon given the frequency f= 3×10^{18} Hz. The frequency is in standard units and we know the relationship (E=hν) between frequency and energy.

Solution:

Step 1: given data

f= 3×10^{18} Hz

Energy =?

Step 2: Write down the formula and rearrange if necessary.

E=hν

Step 2: Put the values in above formula

E=hν

=6.63×10^{-34} J·s ×3×10^{18} Hz=2×10^{-15}J

Result: E= 2×10^{-15}J

Self-Assessment Questions:

1. Ina microwave oven photon of energy1.05×10^{-5} eV is used to heat food.
   (a) Calculate the frequency of the photons.
   (b) Calculate the wavelength of the photons.

25.2 Photoelectric Effect:

- In 1887, Heinrich Hertz observed the photoelectric effect in Germany, discovering that ultraviolet (UV) light shining onto metal surfaces causes the ejection of electrons, which are pushed with potential to reach an anode.
- In 1899, J.J. Thomson in England demonstrated that the incidence of UV light on a metal surface causes the ejection of electrons.
- In 1902, Philip Lenard measured the photoelectric effect quantitatively and concluded that light has electrical properties.

Though these experiments were able to describe phenomenon of photoelectric effect but some situations during experiments were not answered with scientific logic.

In 1905, Albert Einstein was able to resolve the phenomenon. Photoelectric effect is defined as:

When electromagnetic radiation like light is shined on certain metallic materials; electrons are emitted due to absorption of light by the electrons on the surface of material.

25.2.1 Photoelectric Effect Phenomenon:

The photoelectric phenomenon is shown in figure 25.4 in which an evacuated tube containing two electrodes cathode and anode; connected with a variable source voltage V. A monochromatic light source is shined onto cathode through a quartz window and anode (collecting electrode) is connected either positive or negative potential with respect to the cathode. An ammeter is connected with circuit to record the current due to photo-electrons.

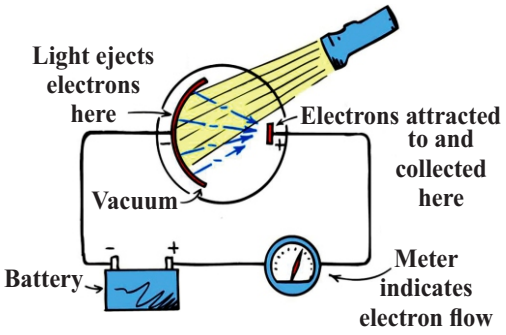


Figure 25.4 Photoelectric Effect

Following observations were made during this experiment:

1. For a constant potential difference between the cathode and anode, the number of electrons emitted from cathode increases with increasing intensity of radiation.
2. For a constant intensity and frequency of incident radiation the photoelectric current varies with the potential difference V between the cathode and anode and reaches a constant value beyond which further increase of potential difference does not affect the photoelectric current, instead, if the anode is made more and more negative with respect of the photo-cathode surface the current decreases. This negative potential (with respect to cathode) of the plate is called ***retarding potential***. As shown in fig: 25.5.
   For a particular value of retarding potential, the photoelectric current becomes zero.
   This potential is called c***ut-off or stopping potential*** V_0 and is measure of maximum kinetic energy of photo-electrons and we can write
   $$K.E_{max} = eV_0$$
   where $K.E_{max}$ (the maximum kinetic energy of the ejected electron.)

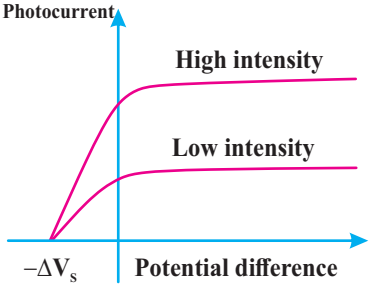
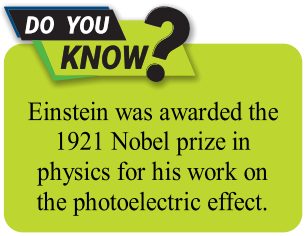


Figure 25.5
graph of photo current vs potential



3. The stopping potential and hence the maximum kinetic energy $K.E_{max}$ of photo-electrons is independent of the intensity of incident radiation and depends only on the frequency ν of radiation.
4. For each substance there exists a characteristic frequency ν_0 such that for radiation with frequency below the photo-electrons are not ejected from the surface. This frequency is called the ***threshold frequency*** and the corresponding wavelength is called ***Threshold wavelength***, λ_0 $(\frac{hc}{\varphi})$
5. The time lag between the incidence of radiation and the emission of a photoelectron is very small, less than 10^{-9}seconds.

25.2.2 Explanation of photo-electric effect on Quantum Theory:

Einstein proposed that radiation energy is not continuously distributed over the wave-front, but the light energy consists of discrete quanta of energy $h\nu$,which penetrates the surface of the cathode, all of its energy is transformed to an electron; it's depicted in figure 25.6. The photon's energy would be associated with its frequency ν, through a proportionality constant known as Planck's constant h, or alternately, using the wavelength λ and the speed of light c:

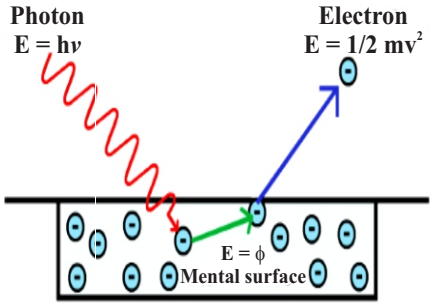


Figure 25.6 photo-electric effect

$$K.E = h\nu = \frac{hc}{\lambda}$$

or the momentum equation:

$$p = \frac{h}{\lambda}$$

When a photon is incident on the surface of a material, some of its energy is spent in making the electron free and the rest appears as kinetic energy of the electron. The electrons at the surface of the material are most loosely bound and require minimum energy for their liberation. This energy is called the ***work function*** ϕ of the material. The maximum kinetic energy of photo electrons ejected from the surface is given by

$$K.Emax = \frac{1}{2}mv^2 = h\nu - \phi \quad \ldots\ldots\ldots\ldots 25.6$$

So, ***work function is the minimum energy that must be supplied to the electron for it to leave the Metal surface. As shown in fig: 25.7.***

Table 25.1
Threshold frequencies value

Metal	Work function ϕ	
Name	*Symbol*	*eV*
Sodium	Na	2.4
Calcium	Ca	2.8
Potassium	K	2.3
Copper	Cu	4.4
Silver	Ag	4.3
Lead	Pb	4.0
Platinum	Pt	6.4
Aluminum	Al	4.28
Gold	Au	5.10
Iron	Fe	4.7
Tungsten	W	4.55

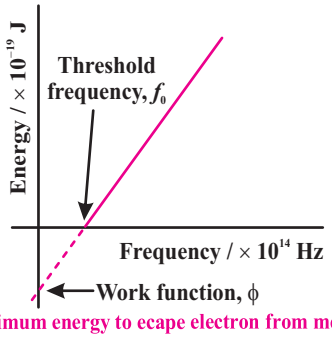


Figure 25.7 a minimum energy to ecape electron from metal surface graph

Consequently there exists a minimum frequency that is independent of the intensity of light, below which electrons cannot be ejected from the metal. This proposal was proved by Millikan when in 1914 he published his results of the voltage required to stop photoelectrons ejecting from a metal surface.

Self-Assessment Questions:

1. How does the photoelectric effect demonstrate the particle nature of light?
2. Calculate the threshold frequency of incident radiation that will cause the emission of photo electrons from the surface of gold.

Worked Example 25.3

The work function for zinc is 4.24 eV. What is the threshold frequency for the ejection of photoelectrons from zinc?

Step 1: When an electron absorbs a photon at threshold frequency, it will just barely have enough energy to overcome the binding forces holding it in the metal and it will emerge with zero kinetic energy.

Step 2: **Write down the formula and rearrange if necessary.**

$$K.Emax = \frac{1}{2}mV^2 = h\nu - \phi$$

Step 3: **First, we determine the frequency in terms of the wavelength.**

$$K.Emax = \frac{1}{2}mV^2 = h\nu - \phi$$

$$0 = h\nu - \varphi_o$$

$$\nu = \frac{\phi}{h} = \frac{4.24 \times 1.6 \times 10^{-19}}{6.63 \times 10^{-34}}$$

$$\nu = 1.02 \times 10^{15}\ Hz$$

Result: $\nu = 1.02 \times 10^{15}\ Hz$

Photoelectric effect in solar cells and photocells:

Photovoltaic cells are made up of p-type and n-type silicon semiconductors. When light falls on the junction between p-type and n-type semiconductors, electrons are emitted according to the photoelectric effect. These electrons are collected in an external circuit to produce an electric current.

- Photocell, a device that uses the photoelectric effect to generate or control a current.
- Photovoltaic cells and photoconductive cells are examples of photoelectric cells.
- Solar cell, a device that converts electromagnetic energy into electrical energy.
- Photovoltaic cells are used to power a wide variety of devices, and are often used in conjunction with a battery that stores the power they produce.

25.3 Compton Effect:

A. H. Compton (1892-1962), aimed short-wavelength light (i.e., x-rays or γ-rays) at various materials, and detected light scattered at various angles. He observed that the scattered light had a slightly longer wavelength than did the incident light, and therefore a slightly lower frequency indicating a loss of energy. He explained this result based on the photon theory as incident photons colliding with electrons.

25.3.1 Particle Model of Photons with Energy and Frequency:

In Compton Effect experiment, there is increase in wavelength of photons, due to their scattering by an electron. The impact results in one of the fundamentals of quantum mechanics, which represents wave and particle properties of light.

Arthur Holly Compton successfully demonstrated that X-rays can be treated as discrete bundles, or quanta, of electromagnetic energy. This concept was later termed "photon" by American physicist Gilbert Lewis.

Photons exhibit both particle-like properties, such as energy and momentum, and wave-like properties, such as frequency and diffraction. The energy of a photon is dependent on its frequency, with lower energy photons having lower frequencies and longer wavelengths, as described by the equation $E = \frac{h\nu}{\lambda}$.

In Compton's experiment, photons collide with free or loosely bound electrons in matter. During these collisions, photons transfer part of their energy and momentum to the electrons, causing the electrons to recoil. This interaction results in the emission of new photons with lower energy and longer wavelengths, a phenomenon known as the Compton shift. The shift in frequency of the scattered photons depends on the energy transferred to the electrons and is independent of the initial wavelength of the incident photons.

25.3.2 Compton Effect Qualitatively:

Compton's Effect, as illustrated in Figure 25.8, can be qualitatively explained as follows:

Initial Interaction:

- A photon, which is a packet of electromagnetic energy, interacts with an electron in a material, typically a target like a metal or graphite.

Scattering Process:

- During the interaction, the photon transfers some of its energy and momentum to the electron. This transfer causes the photon to change its direction and wavelength (or equivalently, its frequency).

Change in Photon Energy:

- The scattered photon emerges with less energy (longer wavelength) than the initial incident photon. The amount of energy lost by the photon is directly related to the energy and momentum gained by the electron.

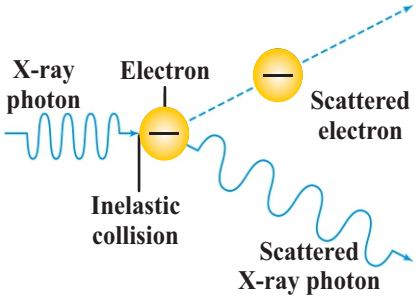


Figure 25.8 Compton Effect

Quantum Nature:

- Compton's Effect cannot be explained using classical wave theory alone. Classical wave theory predicts that light should scatter uniformly without a change in wavelength. However, Compton's observations demonstrated that the scattered light has a shifted wavelength, indicating a particle-like interaction.

Experimental Confirmation:

- Compton conducted experiments where X-rays were targeted at graphite and the scattered X-rays were observed. By measuring the scattering angle and change in wavelength, he confirmed that the results were consistent with the predictions of quantum theory.

Wavelength Shift:

- The wavelength shift observed in Compton scattering is directly proportional to the Compton wavelength, which depends on the mass of the electron and the Planck constant. This relationship provides crucial evidence for the particle nature of photons.

The Compton shift can be evolved as under:

$$\lambda_2 - \lambda_1 = \frac{hc}{E_o}(1 - Cos\theta) = \frac{h}{m_0 c}(1 - Cos\theta)$$

$$\Delta\lambda = \frac{h}{m_0 c}(1 - Cos\theta) \ldots\ldots\ldots\ldots 25.7$$

This equation describes the phenomenon known as Compton Effect. $\Delta\lambda$gives the change in photon wavelength due to scattering with a free electron and it is called **Compton shift.**

It's clear that the Compton Shift is independent of the wavelength of the incident photon and depends on scattering angle.

The term $\lambda_c = \frac{h}{m_0 c} = 2.426 \times 10^{-12} m$, is called Compton Wavelength of the scattering particle i.e., electron.

Self-Assessment Questions:

1. Compare and contrast the photoelectric effect with the Compton effect and the Bohr model of the hydrogen atom in terms of their contributions to the development of quantum theory.
2. How does the Compton Effect demonstrate the particle nature of light?

Worked Example 25.4

A beam of x-rays of wavelength 0.01 Å is incident on a block of graphite. What is the wavelength of the x-rays scattered at an angle of 30°?

Step 1: given data and what to be found

$\theta = 30^o$

$\lambda = 0.01$ Å

Step 2: **Write down the formula and rearrange if necessary.**

$$\Delta\lambda = \frac{h}{m_0 c}(1 - Cos\theta)$$

Step 3: $\Delta\lambda = 2.426 \times 10^{-12}(1 - Cos\ 30^o) = 0.003$ Å

Result: Hence, the final wavelength of these x-rays is $\lambda + \Delta\lambda = 0.01 + 0.003 = 0.013$ Å

25.4 Pair Production:

- Pair production is a phenomenon in quantum mechanics that occurs when a photon with sufficient energy interacts with matter and spontaneously transforms into a particle-antiparticle pair i.e., an electron and a positron, as shown in figure 25.9.

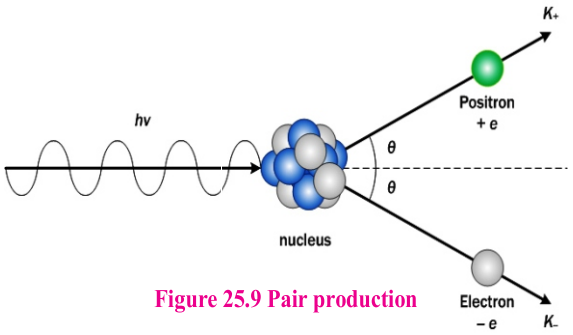


Figure 25.9 Pair production

25.4.1 Conservation laws of Pair-Production:

For pair production to occur, a photon must have a minimum energy of 1.022 MeV. This requirement arises because the rest mass of both an electron and a positron is 0.511 MeV each. Therefore, the combined energy needed to create an electron-positron pair is 0.511 MeV × 2 (1.022 MeV); If the incident photon possesses energy exceeding 1.022 MeV, the surplus energy is distributed as kinetic energy between the electron and the positron.

Therefore, the pair production reaction is given as under:

$$\gamma \rightarrow e^- + e^+ \cong 1.02 \text{ MeV}$$

Facts of pair production process are:

1. The pair production process obeys law of conservation of energy, momentum and electric charge respectively.
2. During collision, the antiparticle of an electron i.e., positron has the same physical properties as electron, except its charge, as both have opposite charge to each other. The sum of charges happens to be zero which is equal to photon before interaction. Therefore, **electric charge is conserved**.
3. The law of conservation of energy is
   (a) $hf = (K.E)_{e^-} + (K.E)_{e^+}$
   (b) $hf = 2m_oc^2 + (K.E)_{e^-} + (K.E)_{e^+}$

DO YOU KNOW?

"Pair production cannot take place in space!"
Pair Production Requires an External Object: Necessary for pair production to occur!

1. Absorbs recoil momentum
2. Conserves energy and momentum simultaneously

DO YOU KNOW?

Antimatter or Antiparticles:
Predicted by British Physicist Paul Dirac, discovered by Carl Anderson

- ✓ Same mass, opposite charge.
- ✓ Every particle has an antiparticle counterpart.
- ✓ Antiparticles have opposite:

Charge:

- ➢ Baryon Number
- ➢ Lepton Number
- ➢ Strangeness

- ✓ Made from anti-quarks (if particles are made from quarks)

25.5 Annihilation of Matter:

Electron-positron annihilation occurs when an electron and a positron collide. The result of the collision is the conversion of the electron and positron to the creation of gamma-ray photons.

25.5.1 Conservation laws of annihilation of matter:

$$e^- + e^+ \rightarrow \gamma + \gamma$$

Each photon has an energy equal to the rest mass of electrons 0.51 MeV. The two photons are produced moving in opposite direction in order to conserve momentum and energy which forbids the creation of a single gamma- ray, as shown in figure 25.10. The charge is also conserved as net charge before and after is zero.

Pair production and Annihilation of matter supports that energy and mass are inter convertible i.e. $E=mc^2$.

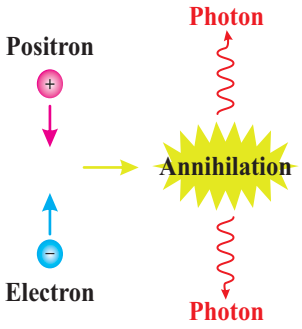


Figure 25.10 Annihilation of Matter

25.6 Wave Nature of Particles:

Classical physics traditionally treated particles and waves as separate and distinct entities.

However, the advent of quantum theory proposed a dual character for radiations, suggesting that they could exhibit characteristics of both waves and particles.

The wave nature of particles implies that particles can behave like waves, displaying properties such as reflection, refraction, interference, diffraction, and other wave-like characteristics. In essence, this duality suggests that matter and radiation can co-exist both particle and wave attributes.

Examples of particles that exhibit this dual nature include:

(i) Matter particles, such as electrons, protons, and neutrons.
(ii) X-rays.
(iii) Photons.
(iv) Electromagnetic radiation.

Particles, like above, can behave like waves. This means they can show wave-like properties, like diffraction and interference. This dual behavior (both wave-like and particle-like) is a fundamental principle of quantum mechanics, which helps us understand the behavior of matter and radiation at a tiny scale.

25.6.1 De Broglie wave:

Louis de Broglie's proposal of wave-particle duality for particles had a profound impact on the understanding of the nature of matter and laid the foundation for the development of quantum mechanics.

De Broglie's idea suggested that both matter and energy could exhibit both particle and wave properties. This unified description was a departure from the classical distinction between particles and waves, providing a more comprehensive understanding of the behavior of particles at the quantum level.

De Broglie's proposal was instrumental in the development of Niels Bohr's model of the atom. It helped explain why electrons in quantized orbits do not radiate energy continuously but exists in stable orbits, as their orbits are analogous to standing waves.

De Broglie Wavelength:

According to de Broglie hypothesis, ***all matter particles like electrons, protons and neutrons in motion are associated with waves.*** These waves are called de Broglie waves or matter waves. The momentum of photon of frequency f is given by

$$p = \frac{hf}{c} = \frac{h}{\lambda} \qquad \text{since } c = \nu\lambda$$

The wavelength of a photon in terms of its momentum is

$$\lambda = \frac{h}{p} \qquad \text{but } p = mv, \text{ then}$$

According to de Broglie, the above equation is completely applicable to matter particles as well. Therefore, for a particle of mass m travelling with speed v, the wavelength is given by

$$\lambda = \frac{h}{mv} \quad \ldots\ldots\ldots \quad 25.8$$

This wavelength of the matter waves is known as de Broglie wavelength. This equation relates the wave character (the ***wavelength***) and the particle character (the ***momentum***) through 25-3, Planck's constant.

Self-Assessment Questions:

1. Calculate the de Broglie wavelength of an electron travelling through a space at a speed of 10^7 m/s. State whether or not these electrons be diffracted by solid materials (atomic spacing in solid materials is 10^{-10} m).
2. How does the De Broglie wavelength relate to the momentum of a particle?

25.6.2 Davisson – Germer Experiment:

De Broglie hypothesis of matter waves was experimentally confirmed by Clinton Davisson and Lester Germer in 1927. They demonstrated that electron beams are diffracted when they fall on crystalline solids. Since crystal can act as a three-dimensional diffraction grating for matter waves, the electron waves incident on crystals are diffracted off in certain specific directions. Figure 25.11 shows a schematic representation of the apparatus for the experiment.

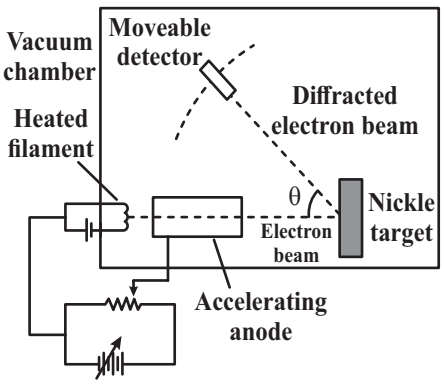


Figure 25.11 Davisson Germer Experiment

The experimental parts are given as under:

1. An electron gun emits electrons via thermionic emission produced by the tungsten filament used in it, i.e., when heated to a specific temperature.
2. Two opposite charged plates known as Electrostatic Particle Accelerator, which accelerates the electrons at a certain potential.

3. The accelerator is enclosed within a cylinder called Collimator which is a narrow passage for the electrons along its axis.
4. The target is a Nickel crystal on which the electron beam is fired normally.
5. When the electrons are scattered from Ni crystal, these are captured by the semicircular movable detector.

Davisson Germer Experiment and de Broglie Relation:

Let us consider an electron of mass (m_0), charge (e) accelerated from rest through a potential V. Then, the kinetic energy $K.E$ of the electron equals to the work done (eV) on it by the electric field:

$K.E = eV$

Now, $K.E = \frac{1}{2}m_0v^2 = \frac{p^2}{2m_0}$

$$p = \sqrt{2m_0K.E} = \sqrt{2m_0eV}$$

Then, the de Broglie wavelength of electron is given by

$\lambda = \frac{h}{p} = \frac{h}{\sqrt{2m_0eV}}$ 25.9

By using $h = 6.626\times10^{-34}$ J-s, $m_0 = 9.11\times10^{-31}$ kg, and $e=1.6\times10^{-19}$ C

we get

$$\lambda = \frac{1.227}{\sqrt{V}} \text{ nm}$$

Where V is the magnitude of accelerating potential in volts.

As, de Broglie wavelength is associated with electrons for V = 54 V from graph 25.12., then we have

$\lambda = 0.167$ nm

DO YOU KNOW?

Bragg's Equation

Electron Diffraction:

- Known: crystal spacing (d) from X-ray diffraction
- Measured: angles (θ) for diffraction peaks
- Calculated: wavelength (λ) of matter waves using: $n\lambda = 2d\sin(90° - \theta/2)$
- Result: $\lambda = 0.165$ nm (wavelength of electron matter waves)

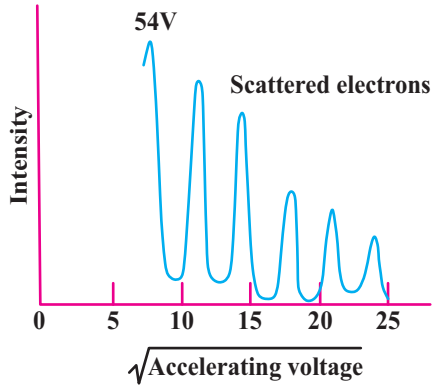


Figure 25 .12 de Broglie wavelength is associated with electrons graph

We conclude that this experiment confirms the wave nature of electrons and the de Broglie relation.

25.6.3 Electron Microscope:

De Broglie hypothesis are used in the field of electron optics. In the electron microscope design, the wave properties of electrons have been utilized with higher resolution, which is giving a great improvement in visualization.

A special type of microscope that uses an electron beam with its wavelike properties to illuminate a specimen able to magnify objects in high resolution (nanometers), which are formed by controlled use of electrons in a vacuum captured on a phosphorescent screen. **Ernst Ruska** (1906-1988), a German engineer and academic professor, built the first Electron Microscope (EM) in 1931, and the same principles behind his prototype still govern modern EMs.

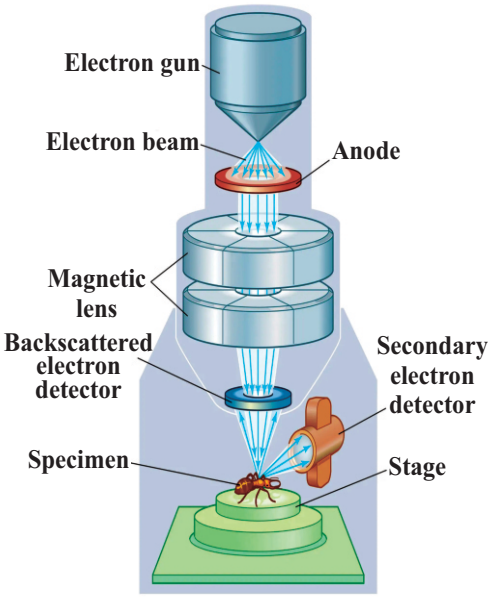


Figure 25.13 Electron Microscope

De Broglie hypothesis of electron have wave paved the way for the development of the electron microscope (EM), which can produce images of much greater magnification than an optical microscope. There are two types of electron microscopes which are transmission electron microscope (TEM), which produces a two-dimensional image, and the scanning electron microscope (SEM), which produces images with a three-dimensional.

Working:

In both types i.e., TEM & SEM, the objective and eyepiece lenses are actually magnetic fields that exert force on the electrons to bring them to a focus. The fields are produced by carefully designed current-carrying coils of wire. EM's measure the intensity of electrons, producing mono chromatic image. Color is often added artificially to highlight. The probe tip of scanning electron microscope, as it is moved horizontally, automatically moves up and down to maintain a constant tunneling current, and this motion is translated into an image of the surface.

Applications:

- Electron microscopes are used to investigate the ultra-structure of a wide range of biological and inorganic specimens including microorganisms, cells, large molecules, biopsy samples, metals, and crystals.
- Industrially, electron microscopes are often used for quality control and failure analysis.

DO YOU KNOW?

Electron microscopes

As the wavelength of an electron can be up to 100,000 times shorter than that of visible light photons, electron microscopes have a higher resolving power than light microscopes and can reveal the structure of smaller objects.

- Modern electron microscopes produce electron micrographs using specialized digital cameras and frame grabbers to capture the images.
- The advancement of microbiology is significantly indebted to the electron microscope, which has revolutionized our understanding of microorganisms such as bacteria, viruses, and other pathogens, thereby greatly enhancing the effectiveness of disease treatments.

25.6.4 Role of Electron Microscope:

1. **Higher Resolution:** Electron microscope offers significantly higher resolution compared to optical microscopes, allowing observation at the nanometer scale due to the shorter wavelength of electrons.
2. **Transmission Electron Microscope (TEM):** TEMs study internal structures of thin specimens, producing detailed images of cells, organelles, and crystalline structures. Valuable in biology, materials science, and nanotechnology.
3. **Scanning Electron Microscope (SEM):** SEMs provide 3D surface images by scanning specimens and detecting emitted secondary electrons. Widely used in biology, geology, and materials science for surface analysis.
4. **Energy- Dispersive X-ray Spectroscopy (EDS):** With EDS detectors, electron microscopes analyze elemental composition by detecting X-rays emitted when high-energy electrons interact with the sample.
5. **Materials Science:** Crucial for studying the microstructure of materials, aiding understanding of relationships between microstructure and properties in metals, ceramics, and polymers.
6. **Nanotechnology:** Essential for imaging and characterizing nano-materials, contributing to the development of new materials and devices through observation and manipulation of nano-particles and nanostructures.

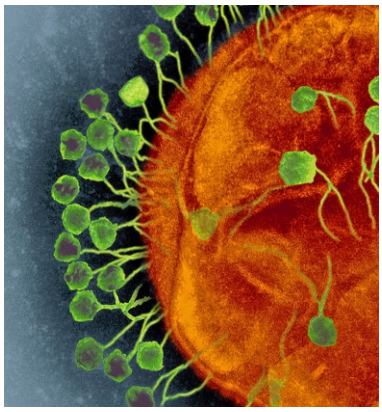


Figure 25.14 (a)
Viruses attacking a cell of bacterium (×50000) by TEM



Figure 25.14 (b)
Viruses attacking a cell of bacterium (×35000) by SEM

7. **Advancements in Medicine:** Contributes to medical research by providing insights into the structure of viruses, bacteria, and cellular organelles, aiding understanding of diseases and therapeutic development.
8. **Quality Control in Industry:** Used across industries for quality control and failure analysis, helping identify defects, analyze material composition, and ensure product integrity at the microscopic level.

25.7 Uncertainty Principle:

In 1927, Werner Heisenberg suggested a principle that applies to measure the properties of quantum-sized objects (e.g., atomic and sub-atomic particles).

The uncertainty principle mainly applied to experiments in physics labs, however some real-world effects are given. For example, in the sun's core, it is the uncertainty in the position of two hydrogen nuclei (protons) that allows for there to be a chance that they will overlap and fuse together. This fusing releases a tremendous amount of energy, which we receive as light. A pretty significant effect on our everyday lives.

25.7.1 Heisenberg Uncertainty Principle:

It states that it is impossible to measure simultaneously, certain pairs of properties of a subatomic particle, such as its position and momentum, with absolute certainty.

In other words the more precisely we know the position of a particle, the less precisely be able to know its momentum, and vice versa.

This principle develops from the wave-particle duality of particles in quantum physics, where particles can exhibit both wave-like and particle-like behaviors.

The formal inequality relating the standard deviation of position Δx and the standard deviation of momentum Δp are given here:

$$\Delta x.\Delta p \geq \frac{h}{4\pi}$$

$$\Delta x.\Delta p \geq \frac{\hbar}{2} \text{.....25.10}$$

ħ (pronounced "h-bar") is the reduced Planck constant

Heisenberg uncertainty principle can also be applied to other pairs of complementary quantum properties, such as energy and time

$$\Delta E.\Delta t \geq h/2\pi \quad \text{..... ..... 25.11}$$

Here,ΔE and Δt represents the uncertainties in the energy and time respectively.

The Uncertainty Principle has profound consequences for understanding quantum world. It means that at the subatomic level, there are inherent limits to how precisely we can measure certain properties of particles. This is not due to restrictions in our measuring instruments but it is a fundamental aspect of the quantum nature of the universe.

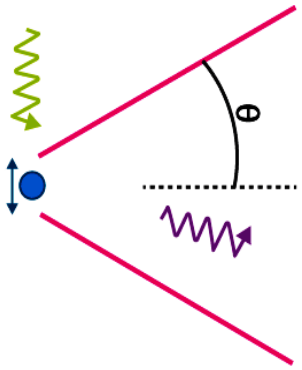


Figure 25.16 Uncertainty Principle

DO YOU KNOW?

Interesting Facts about Planck's Constant (h):

Universal constant: Appears in many areas of physics, like quantum mechanics, thermodynamics, and electromagnetism.

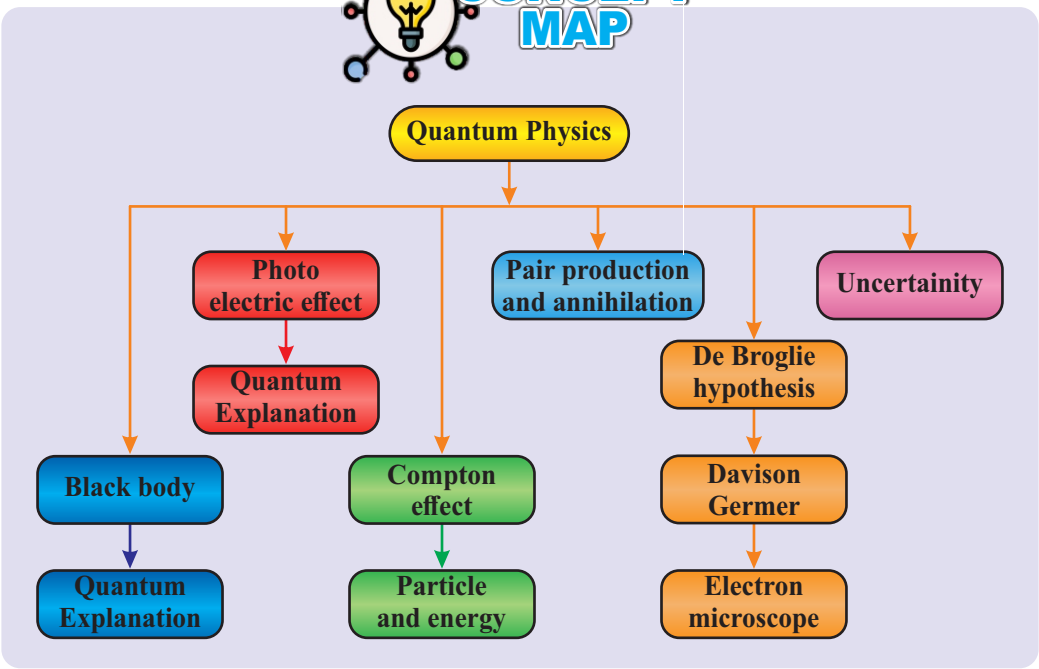
Tiny but mighty: Very small value (approximately 6.626 x 10^-34 J s), but has a huge impact on our understanding of the universe.

Fundamental unit: Used to define the Planck units, a set of fundamental units for length, time, mass, and energy.



- ✓ A black body absorbs all incident radiation and releases it in a continuous spectrum of colors, depending upon its temperature.
- ✓ The energy per unit volume per unit wavelength of black body radiation is directly proportional to the square of the wavelength and increases indefinitely as the wavelength becomes smaller.
- ✓ Different atoms and molecules can emit or absorb energy in discrete quantities only.
- ✓ When electromagnetic radiation like light is shinned on certain metallic materials; electrons are emitted due to absorption of light by the electrons on the surface of material.
- ✓ Work function, is the minimum energy that must be supplied to the electron for it to leave the metal.
- ✓ For each substance there exists a characteristic frequency f_0 such that for radiation with frequency below the photo-electrons are not ejected from the surface.
- ✓ Compton scattering, refers to X-ray or gamma-ray photons undergo a change in wavelength and momentum when they collide with charged particles, typically electrons.
- ✓ Change in wavelength gives the change in photon wavelength due to scattering with a free electron.
- ✓ When the photon interacts with the strong electric field around the nucleus, it undergoes a change of state and is transformed into two particles (creating matter from energy):
- ✓ Electron-positron annihilation occurs when an electron and a positron (the antiparticle of the electron) collide.
- ✓ All matter particles like electrons, protons and neutrons in motion are associated with waves.
- ✓ A microscope which uses a beam of accelerated electrons as a source of lighting.
- ✓ It is impossible to measure simultaneously all quantum properties of a particle with equal accuracy.

CONCEPT MAP



EXERCISE

Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. If h is the Planck constant, then $\hbar$ is:
   (a) $2\pi h$ (b) $2h$ (c) $h/2$ (d) $h/2\pi$ (e) $2h/\pi$
2. In a photoelectric effect experiment the stopping potential is:
   (a) the energy required to remove an electron from the sample
   (b) the kinetic energy of the most energetic electron ejected
   (c) the potential energy of the most energetic electron ejected
   (d) the electric potential that causes the electron current to vanish
3. The work function for a certain sample is 2.3 eV. The stopping potential for electrons ejected from the sample by 7.0×10^{14}-Hz electromagnetic radiation is:
   (a) 0 (b) 0.60 V (c) 2.3V (d) 2.9V (e) 5.2V
4. In Compton scattering from stationary particles the maximum change in wavelength can be made smaller by using:
   (a) higher frequency radiation (b) lower frequency radiation
   (c) more massive particles (d) less massive particles

5. Of the following, Compton scattering from electrons is most easily observed for:
   (a) infrared light (b) visible light
   (c) ultraviolet light (d) X-rays
6. In the Compton scattering from stationary electrons the largest change in wavelength occurs when the photon is scattered through
   (a).0° (b) 45° (c) 90° (d) 180°
7. A free electron has a momentum of 5.0×10^{-24} kg · m/s. The wavelength of its wave function is:
   (a) 1.3×10^{-8}m (b) 1.3×10^{-10} m
   (c) 2.1×10^{-11} m (d) 2.1×10^{-13} m
8. In Photoelectric Effect, when light fall on the surface of metal, the material should emit
   (a) Electrons (b) Protons
   (c) Positrons (d) Neutrons
9. Threshold frequency is defined as the ..... frequency of incident light which can cause photo electric emission
   (a) Maximum (b) Minimum
   (c) Average (d) highest
10. The amount of energy which is necessary to start photo electric emission is called:
   (a) Maximum (b) Average
   (c) Minimum (d) Littlest

Section (B): CRQs (Short Answered Questions):

1. Differentiate between wave and particle.
2. Is it possible for the de Broglie wavelength of a particle?
3. Estimated Broglie wavelength of a cricket ball on the pitch?
4. Differentiate between the continuous and discrete emission of radiation?
5. What is threshold frequency?
6. How has the photoelectric effect been applied in real-world technologies or devices, and what are its practical implications?
7. What are the advantages of an electron microscope over an optical microscope?
8. Is it possible to create only an electron through matter and photon interaction?
9. Give construction of electron microscope?
10. Elaborate the particle nature of electromagnetic radiation.

Section (C): ERQs (Long Answered Questions):

1. Describe how energy is distributed over the wavelength range for different temperatures.
2. Explain the particle model of light in terms of photons with particular energy and frequency.
3. Describe conservation laws in pair production and annihilation of matter.
4. Describe Compton's effect qualitatively.
5. Explain how the very short wavelength of electrons, and the ability to use electrons and magnetic fields to focus them, allows electron microscope to achieve very high resolution.
6. Describe the impact of de Broglie proposal that any kind of particle has both wave and particle properties.
7. Describe the confirmation of de Broglie proposal by Davisson and Germer experiment in which the diffraction of electrons by the surface layers of a crystal lattice was observed.

Section (D): Numerical:

1. The Sun's surface temperature is 5700 K.
   (i) How much power is radiated by the Sun?
   (ii) Given that the distance to Earth is about 200 Sun radii, what is the maximum power possible from one square kilometer solar energy installation?
   (iii) What is the wavelength of maximum intensity of solar radiation?
   (5.98×10^{7} W/m^2, 3.6×10^{26} Watts, 5.1×10^{-7} m)
2. The temperature of your skin is approximately 32 °C. What is the wavelength at which the peak occurs in the radiation emitted from your skin? **(9.05×10^{-5} m)**
3. An FM radio transmitter has a power output of 100 kW and operates at a frequency of 94 MHz, How many photons per second does the transmitter emit?
   (6.23×10^{-26} J, 1.61×10^{30} s^{-1})
4. A light source of wavelength illuminates a metal and ejects photoelectrons with a maximum kinetic energy of 1.0 *eV*. A second light source with half the wavelength of the first ejects photoelectrons with a maximum kinetic energy of 4.0 eV. Determine the work function of the metal. **(2 eV)**
5. A 430 nm violet light is an incident on a calcium photo electrode with a work function of 2.71 eV. Find the energy of the incident photons and the maximum kinetic energy of ejected electrons. **(2,88 eV, 0.17 eV)**
6. Cut-off frequency for the photoelectric effect in some materials is 8×10^{13}Hz. When the incident light has a frequency of 1.2×10^{14} Hz, the stopping potential is measured as – 0.16 V. Estimate a value of Planck's constant from these data and determine the percentage error of your estimation. **(5.17 eV, 1.30 V)**

7. The work function of some metals is listed below. The number of metals which will show photoelectric effect when light of 300 nm wavelength falls on the metal is:

Metal	Li	Na	K	Mg	Cu	Ag	Fe	Pt	W
φ in eV	2.4	2.3	2.2	3.7	4.8	4.3	4.7	6.3	4.75

(4 Metals)

8. X-rays with an energy of 300 keV undergo Compton scattering with a target. If the scattered X-rays are detected at 30° relative to the incident X-rays, determine the Compton shift at this angle, the energy of the scattered X-rays, and the energy of the recoiling electron. **(0.35 pm, 278keV, 22keV)**

9. A photon with a wavelength of 6.0×10^{-12} m collides with an electron. After the collision the photon wavelength is found to have been changed by exactly one (Compton Wavelength is 2.43×10^{-12} m).
   (i) What is the photon's wavelength after collision?
   (ii) Through what angle has been deflected in this collision?
   (iii) What is the angle for the electron after the collision?
   (iv) What is the electron's kinetic energy, in eV, after collision?

   (8.4×10^{-12} m, 90°, < 90°, 5.9×10^{4} eV)

10. Find the de Broglie wavelength of an electron in the ground state of hydrogen.

    (3.324°A)

11. Determine the minimum uncertainties in the positions of the following objects if their speeds are known with a precision of 1.0×10^{-3}m/s: (a) an electron and (b) a bowling ball of mass 6.0 kg. **(5.8cm, 33 m)**

Unit 26 Atomic Physics

Teaching Periods 12 **Weightage %** 08



The "electron cloud" model was proposed by Erwin Schrödinger in 1926, as part of his quantum mechanical theory. Schrödinger's model described the electrons in an atom as a cloud of probability, where the density of the cloud represents the likelihood of finding an electron within a given region. This replaced the earlier planetary model, which had electrons in fixed orbits.

In this unit student should be able to:

- Describe and explain the origin of different types of optical spectra.
- Show an understanding of the existence of discrete electron energy levels in isolated atoms (e.g. atomic hydrogen) and deduce how this leads to spectral lines.
- Explain how the uniqueness of the spectra of elements can be used to identify an element.
- Describe Bohr's postulates of Hydrogen atom
- Derive an expression for quantized radii
- Explain hydrogen atom in terms of energy levels on the basis of Bohr Model
- Determine the ionization energy and various excitation energies of an atom using energy level diagram
- Illustrate the significance of the hydrogen spectrum in the development of Bohr's model of the atom.
- Derive $\frac{1}{\lambda} = RH\left[\frac{1}{P^2} - \frac{1}{n^2}\right]$.
- Solve problems using $\frac{1}{\lambda} = RH\left[\frac{1}{P^2} - \frac{1}{n^2}\right]$.
- Describe inner shell transitions
- Explain production and characteristics of X-rays based on inner shell transition
- Describe properties and uses of X-rays
- Explain the terms spontaneous emission, stimulated emission, Meta stable states, population inversion and laser action.
- Describe the structure and purpose of the main components of a He-Ne gas laser.
- Identify the useful properties of laser light and give some examples of their uses
- Identify the measures requirement for safe handling of lasers

Atomic Spectra:

Atomic spectra deal with the spectrum of frequencies of electromagnetic radiation emitted or absorbed by an electron during transitions between different levels of energy within an atom.

Electrons present in an atom have discrete and specific energies. There are more energy states in an atom than there are electrons. The rainbows are the most common example of atomic spectra, even if it may seem as a continuous pattern; it has black lines that represent the absorption spectra of the sun. The spectrum is also used in astronomy to identify the composition of stars.

26.1.1: Optical Spectra:

An optical spectrum is obtained, when light passes through a prism or a diffraction-grating. This spectrum can be obtained from the intensity and wavelength of the radiation.

The spectrum may be observed visually in the limited wavelength region to which the eye is sensitive; it may be focused on a photographic plate. The spectra obtained from radiating bodies are called emission spectra and are classified into three types

(i) **Continuous spectra:** These are emitted by solids, liquids and dense opaque gases at high temperatures. The spectrum of the sun, or of a black body, is a continuous spectrum. Gases at low pressures emit line spectra.

(ii) **Band spectra:** These are associated with similar changes in the molecules under high pressure.

(iii) **Line spectra:** These have their origin in the energy changes which take place in the atoms of a gas, while

An absorption spectrum is used to analyze the substance and this spectrum is obtained by passing white light from a continuous source through a gas or a dilute solution of the element being analyzed. The absorption spectrum consists of a series of dark lines superimposed on the continuous spectrum of the light source as shown in figure 26.2.

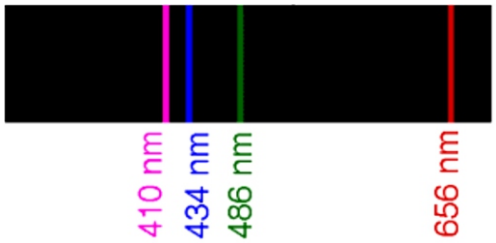


Figure 26.1 Types of emission spectra

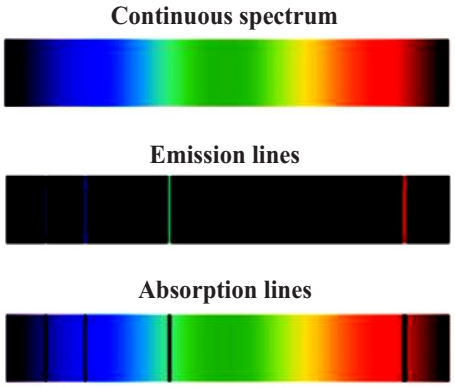


Figure 26.2 hydrogen absorption spectrum

26.1.2: Spectral lines:

A line of particular frequency or wavelength emitted or absorbed by atoms is called spectral line.

The four visible lines occur at the wavelengths 656 nm, 486 nm, 434 nm, and 410 nm. The complete set of lines is called the Balmer series. The wavelengths of these lines can be described by the following equation,

$$\frac{1}{\lambda} = R_H\left(\frac{1}{2^2} - \frac{1}{n^2}\right); n = 3,4,5, \ldots..26.1$$

1. $Balmer\ Series\ \frac{1}{\lambda} = R_H\left(\frac{1}{2^2} - \frac{1}{n^2}\right)$
2. $Paschen\ Series\ \frac{1}{\lambda} = R_H\left(\frac{1}{3^2} - \frac{1}{n^2}\right)$
3. $Lyman\ Series\ \frac{1}{\lambda} = R_H\left(\frac{1}{1^2} - \frac{1}{n^2}\right)$
4. $Brackett\ Series\ \frac{1}{\lambda} = R_H\left(\frac{1}{4^2} - \frac{1}{n^2}\right)$
5. $PFund\ Series\ \frac{1}{\lambda} = R_H\left(\frac{1}{5^2} - \frac{1}{n^2}\right)$

where R_H is a constant called the Rydberg constant. The above equation and other lines in the spectrum of hydrogen atom (Lyman, Paschen, Brackett & Pfund Series) will be disused in the section

DO YOU KNOW?

Between 1860 and 1885, scientists collected data on atomic emissions using spectroscopy. In 1885, Johann Jacob Balmer discovered an equation that accurately predicted the wavelengths of hydrogen's four visible emission lines: Ha (red), Hb (green), Hg (blue), and Hd (violet),

26.2.1.3 The spectrum of hydrogen atom:

The shape of the spectrum proposed by Max Planck in 1900, that the energy of a given mode of vibration depended upon its frequency in accordance with the relationship.

$$E_n = nhv; n = 1,2,3, \ldots \ldots \ldots 26.2$$

This relation shows that the energy associated with each frequency for different transition has the discrete value.

DO YOU KNOW?

Early solar spectrum studies revealed unknown lines, leading to the discovery of helium, named after the Greek word for Sun, Helios. Since then, analyzing starlight has only detected elements found on Earth, suggesting a universal consistency of elements.

26.1.3: Identification of Elements by Uniqueness of Spectra:

Each element has its own unique line spectrum which is referred as the "fingerprint" of a particular element.

The spectra for each element are unique because each element contains different numbers of electrons and different energy levels. The absorption spectrum of an element has many practical applications. For example, various absorption lines observed in the solar spectrum have been used to identify elements in the solar atmosphere.

Self-Assessment Questions:

1. Why are some spectral lines brighter than others?
2. Why different elements have the different spectra?
3. Differentiate between emission and absorption spectra.

26.2 Bohr's Model and its Postulates:

In 1913, Niels Bohr introduced atomic model in order to give quantitative determination of frequency emitted during de-excitation of an electron in Hydrogen -atom.

The discovery of the electron in the late 19th century marked the beginning of a new era in scientific research, helping physicists understand the structure and nature of atoms

Niels Bohr had a remarkable idea that explained the atom of hydrogen-consist of a proton and an electron revolving around the nucleus as shown in figure 26.3.

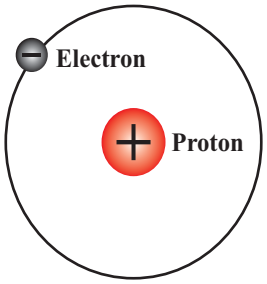


Figure 26.3 hydrogen atom

The following are the postulates of Bohr

Postulates I:

The electron in a hydrogen atom orbits the nucleus,

$$\mathbf{F}_C = \frac{mx\,v^2}{r} = \mathbf{F} = k \times \frac{q_1\,x\,q_2}{r^2}$$

Postulates II:

The magnitude of the angular momentum L of the electron in its orbit is equal to the integral multiple of $\frac{h}{2\pi}$ i.e.,

$L = n\frac{h}{2\pi}$, n= 1,2,3,4,…….

where h is called Plank's constant and n is positive integer (quantum number).

Postulates III:

Only certain orbits are stable in which electrons are revolving and these orbits are called stationary states. The atom emits radiation (photon) when the electron makes a transition from a higher energy state (E_n) to the lower energy state (E_p).

$$h\nu = E_n - E_p \ldots\ldots\ldots\ldots 26.3$$

where ν is the frequency of emitted photon.

DO YOU KNOW?

Limitations of Bohr's Model are:

1. Failed to explain Zeeman Effect.
2. Failed to explain Stark Effect.
3. Could not explain spectra obtained from larger atoms.

These phenomena were later explained by Schrodinger Atomic Model

26.2.2 Bohr's radius:

The electron revolving around the nucleus is in uniform circular motion and thus experiences a centripetal force. Here centripetal force is provided by electrostatic force between electron and proton.

$$F_c = \frac{mv^2}{r}(i)$$

where V is the speed of electron. The magnitude of electrostatic force $\left(F_e = \frac{kq_1q_2}{r^2}\right)$between the electron $(q_1 = -e)$ and the proton $(q_1 = e)$ separated by the orbital radius r is given as.

$$F_e = \frac{ke^2}{r^2} \qquad (ii)$$

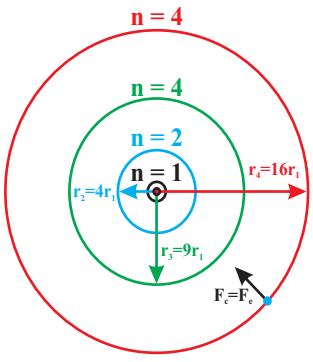


Figure 25.4 quantized radii for hydrogen atom

where $k = 9.0 \times 10^9 Nm^2/C^2$is called Coulomb's constant.
The electron can only move in a particular orbit if the above two forces are balanced each other. Comparing equation (i) and (ii)

$$\frac{ke^2}{r^2} = \frac{mV^2}{r} \qquad (iii)$$

From Bohr's postulate (2),

$L = mVr = n\frac{h}{2\pi}$

Substituting the value of $V = \frac{nh}{2\pi mr}$

$$\frac{ke^2}{r^2} = \frac{m}{r}\left(\frac{nh}{2\pi mr}\right)^2$$

Where,

$k = \frac{1}{4\pi\varepsilon_0}$

$h = 6.63 \times 10^{-34} js$

$m = 9.1 \times 10^{-31} kg$

$\varepsilon_0 = 8.85 \times 10^{-12} C^2/Nm^2$

$r = \frac{n^2\varepsilon_0 h^2}{\pi me^2} \qquad (IV)$

$r = r_0 n^2 \qquad (v)$

where $r_0 = 5.29 \times 10^{-11} m = 0.53A^0$ is called Bohr's first orbit.
As the values of *r* is depending only on the principle quantum number *n*. Therefore, the radii have the quantized values as shown in figure 26.4.

$$r_n = r_1 n^2 \qquad \ldots\ldots\ldots 26.4$$

26.2.3: Orbital Energy of hydrogen atom:

The allowed energy levels and quantitative values of the emission wavelengths of the hydrogen atom can be calculated from the postulate (3) suggests the qualitative existence of a characteristic discrete emission spectrum and corresponding absorption spectrum for hydrogen. The electron has kinetic energy ($K = \frac{1}{2}mV^2$) and electric potential energy ($U = -\frac{kq_1q_2}{r}$). The total energy will be

$$E = K + U$$

$$E = \frac{1}{2}mV^2 - \frac{ke^2}{r} \ldots\ldots\ldots 26.5$$

Putting the value of $mV^2 = \frac{ke^2}{r}$ from equation (III)

$$E = \frac{ke^2}{2r} - \frac{ke^2}{r}$$

$$E = -\frac{ke^2}{2r}$$

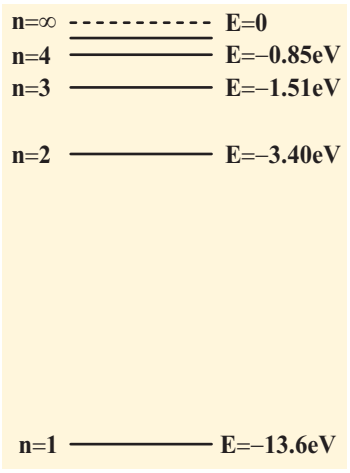


Figure 26.5
various excitation levels

Substitute the value of r from equation(IV)

$$E = -\frac{me^4}{8\varepsilon_0{}^2h^2}\left(\frac{1}{n^2}\right) \qquad (26.6)$$

The numerical value of this energy will be

$$E_n = \frac{-13.6}{n^2}eV \qquad (26.7)$$

The energy level diagram (see figure 26.5) shows the various -excitation levels.

26.2.4 Excitation & Ionization Energy:

Excitation:

The process of transferring energy to an atom or molecule, raising it to a higher energy state, often resulting in the emission of photons as it returns to its ground state.

Ionization Energy:

The minimum energy required to remove an electron from an atom or molecule, resulting in the formation of an ion from figure 26.5 the energy levels corresponding to various of n as given inequation $E_n = \frac{-13.6}{n^2}eV$ are mentioned. The lowest level for *n=1* is the ground state of hydrogen atom and higher levels correspond to excited states. The energy for the greatest value of $n = \infty$ become zero and for any energy $E = 0$ in which the electron and proton do not bound together is called ionization energy. Therefore, the energy E = 13.6 eV is required to ionize the hydrogen atom.

26.2.5 The spectrum of Hydrogen atom:

The electron in a hydrogen atom can jump between quantized energy levels by emitting or absorbing Photon for some different values of wavelengths. Any such wavelength is often called a line spectrum which can be absorption or emission lines. The lines for hydrogen are said to be grouped into series, according to the level at which upward jumps start and downward jumps end. The formula for these series corresponding to the different wavelengths can be obtained from equation

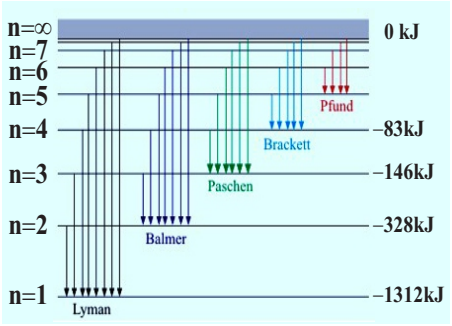


Figure 26.6 Atomic spectrum of hydrogen

$$E = -\frac{me^4}{8\varepsilon_0{}^2h^2}\left(\frac{1}{n^2}\right)$$

The frequency of the photon emitted when the electron makes a transition from an orbit to an inner orbit as stated in postulate **-III**.

$\nu = \frac{(E_n - E_p)}{h}$ using equation (6)

$$\nu = \frac{me^4}{8\varepsilon_0{}^2 h^3}\left(\frac{1}{P^2} - \frac{1}{n^2}\right)$$

To express above equation in terms of wavelength, it is convenient to use

$c = \nu\lambda$

$$\frac{1}{\lambda} = \frac{me^4}{8c\varepsilon_0{}^2 h^3}\left(\frac{1}{P^2} - \frac{1}{n^2}\right)$$

where $R_H = \frac{me^4}{8c\varepsilon_0{}^2 h^3} = 1.097 \times 10^7 m^{-1}$, is called Rydberg constant.

$$\frac{1}{\lambda} = R_H\left(\frac{1}{P^2} - \frac{1}{n^2}\right) \qquad (26.8)$$

The possible series associated the different wavelengths can be explained from equation (8). The emission and absorption lines for all possible jumps up from the $n = 1$ level and down to the $n = 1$ level are said to be in the Lyman series. Similarly, other series are illustrated in the figure 26.6.

Worked Example 26.1

The electron in a hydrogen atom makes a transition from the n =3 energy level to the ground level (n = 1). Find the wavelength and frequency of the emitted photon.

Solution:

Step 1: Write down the known quantities and quantities to be found.

n =3

p = 1

$$R_H = \frac{me^4}{8c\varepsilon_0{}^2 h^3} = 1.097 \times 10^7 m^{-1}$$

Step 2: Write down the formula and rearrange if necessary.

$$\frac{1}{\lambda} = R_H\left(\frac{1}{P^2} - \frac{1}{n^2}\right)$$

$$c = \nu\lambda$$

Step 3: Put the values in formula and calculate.

$$\frac{1}{\lambda} = R_H\left(\frac{1}{1^2} - \frac{1}{3^2}\right) = \frac{8}{9}R_H$$

$$\frac{1}{\lambda} = 0.975 \times 10^7 m^{-1}$$

$$\lambda = 102.56\, nm$$

$$\nu = \frac{c}{\lambda} = \frac{3.0 \times 10^8 m/s}{102.56 \times 10^{-9} m}$$

$$\nu = 2.925 \times 10^{15} Hz$$

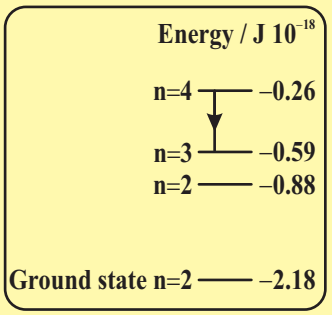
Worked Example 26.2

The lowest energy levels of a mercury atom are shown below. The diagram is not to scale. Calculate the frequency of an emitted photon due to a transition, shown by an arrow, from level n = 4 to level n = 3. Which transition would cause the emission of a photon of a longer wavelength than that emi1ted in the transition from level n = 4 to level n = 3?

Solution:

Step 1: Write down the known quantities and quantities to be found.

$$E_4 = -0.26 \times 10^{-18} j$$
$$E_3 = -0.59 \times 10^{-18} j$$
$$h = 6.63 \times 10^{-34} js$$



Step 2: Write down the formula and rearrange if necessary.

$$h\nu = E_4 - E_3$$
$$c = \nu\lambda$$

Step 3: Put the values in formula and calculate.

$$\nu = \frac{-0.26 \times 10^{-18} + 0.59 \times 10^{-18} j}{6.63 \times 10^{-34} js}$$
$$\nu = 4.9773 \times 10^{14} Hz$$
$$\lambda = \frac{c}{\nu} = \frac{3.0 \times 10^{8} m/s}{4.9773 \times 10^{14} Hz}$$
$$\lambda = 6.027 \times 10^{-7} m$$
$$h\nu = E_n - E_p = \Delta E$$
$$\lambda = \frac{hc}{\Delta E}$$

From this relation, it clear that the wavelength will increase for lowest value of ΔE. The longer wavelength can be calculated from the transition n = 3 to n = 2.

$$\lambda = \frac{hc}{\Delta E} = \frac{6.63 \times 10^{-34} js \times 3.0 \times 10^{8} m/s}{(0.88 - 0.59) \times 10^{-18} j}$$
$$\lambda = 6.858 \times 10^{-7} m$$

Self-Assessment Questions:

1. A mercury atom de-excites from its 4.9 eV energy level to the ground state. Calculate the wavelength of the photon released.
2. What is the ratio of the shortest wavelength of the Balmer series to the shortest wavelength of the Lyman series?

26.3 X-Rays:

These are powerful electromagnetic rays of higher energy and can pass through most objects, including the body. Most of them have a wavelength ranging from 0.01 to 10 nanometres, corresponding to frequencies 3×10^{19} Hz to 3×10^{16} Hz. X-rays are used to generate images of tissues and structures inside the body.

DO YOU KNOW?

The discovery of X-rays, was made by the German physicist Wilhelm Conrad Roentgen in 1895. He named them X-rays to highlight the fact that their nature was unknown.

26.3.1 Inner Shell Transitions:

The passage of electron from one energy level to another is called electron transition. Bohr Theory showed that the principal quantum number is used to determine the energy and an orbital radius of electron. The inner shells, designated K, L, M, etc., correspond to 1,2,3, respectively, the K shell having the greatest binding energy and being closest to the nucleus. The other quantum numbers have a relatively small effect on energy and cause the shells (other than the K shell) to be split into subshells. When the transition of electrons from the outer shell to the inner shell occurred then the electromagnetic radiation corresponding to the characteristic energy generated as shown in figure: 26.7.

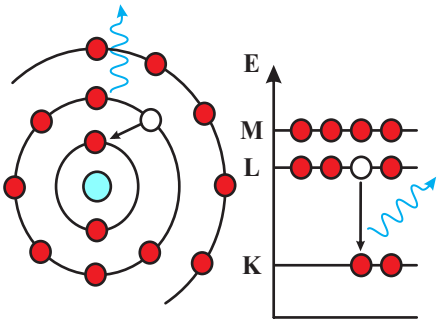


Figure: 26.7. Inner Shell Trasition

26.3.2 Production and characteristics of X-rays:

In 1895, Roentgen found that some photographic plates, kept carefully wrapped in his laboratory, had become fogged. Instead of merely throwing them aside he set out to find the cause of the fogging. He traced it to a gas-discharge tube, which he was using with a low pressure and high voltage. This tube appeared to emit a radiation that could penetrate paper, wood, glass, rubber, and even aluminum a centimeter and a half thick. Roentgen could not find out whether the radiation was a stream of particles or a train of waves. Newton had the same difficulty with light and he decided to call it X-rays.

DO YOU KNOW?

Roentgen's first X-ray was of his ***wife's left hand***, on which was her wedding ring. However, when she saw the image, she was not impressed. She said, "I have seen my death!



X-rays as electromagnetic waves, but of much shorter wavelength: about 0.1nm to 10nm. They are produced when fast electrons, or cathode rays, strike a target, such as the walls or anode of a low-pressure discharge tube. In a modern X-ray tube there is no gas, or as little as high-vacuum technique can achieve: the pressure is about 10^{-5} mm Hg. The electrons are provided by thermionic emission from a white-hot tungsten filament.

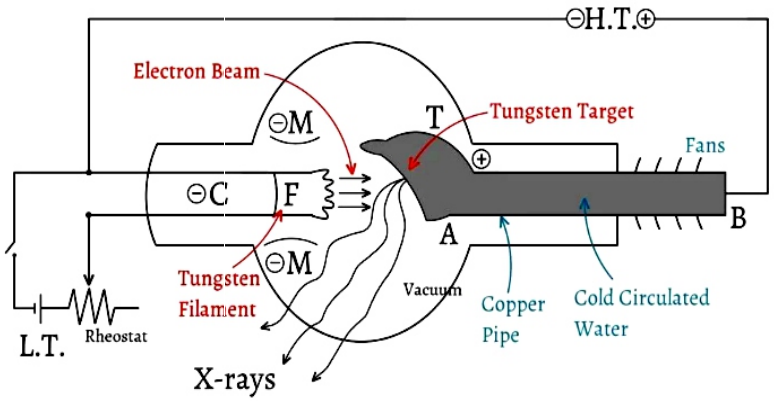


Figure: 26.8. Production of X-Rays

In Fig. 26.8 F is the filament and T is the target, or anode. Because there is so little gas, the electrons on their way to the anode do not lose any noticeable amount of their energy in ionizing atoms. From the a.c. supply, transformers provide about 10 volts (L.T) for heating the filament, and about 100 kV (H.T) for accelerating the electrons. On the half cycles when the target is positive, the electrons bombard it, and generate X-rays. On the half-cycles when the target is negative, nothing happens at all—there is too little gas in the tube for it to break down. Thus the tube acts, in effect, as its own rectifier providing pulses of direct current between target and filament. The heat generated at the target by the electronic bombardment is so enormous that the target must be cooled artificially.

DO YOU KNOW?

L.T (Low Tension): This refers to low voltage.
H.T (High Tension): This refers to high voltage.
Spine needs the longest exposure to X-ray, it exposed for 0.20 second to get proper image.

In an X-ray tube, very energetic electrons bombard on the atoms in a metal target, the atom will be left in an excited state with a hole in the electron shell. This process is illustrated in Fig. 26.7 for inner shell transition. When this hole is filled by an electron from an outer shell, an x-ray photon with energy equal to the difference in the electron energy levels is produced. The energy of the x-ray photon is characteristic of the target metal. The different characteristic lines occurred due to the electron transitions between different energy levels as shown in figure 26.9. Electron transitions in an atom, which produce the K_α, K_β and L_α characteristic x-rays.

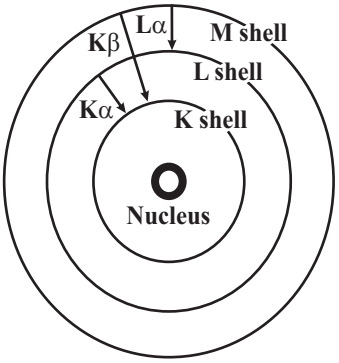


Figure: 26.9.
Electron Transition in atom

26.3.3 Properties and Uses of X-rays:

When X-rays strike some minerals, such as zinc sulphide, they make them fluoresce. If a human body part is placed between an X-ray tube and a fluorescent screen, the shadows of

its bones can be seen on the screen, because they absorb X-rays more than flesh does. Unusual objects, such as swallowed safety-pins, if they are dense enough, can also be located. X-ray photographs can likewise be taken, with the plate in place of the screen. In this way cracks and flaws can be detected in metal castings. The characteristics lines are most useful in x-ray diffraction for studying the crystallographic structures.

Self-Assessment Questions:

1. On what factors the shortest wavelength of X-rays emitted from tube depends?
2. What determines the characteristic wavelength emitted by an X-ray tube?

26.4.1 LASER [*Light Amplification by Stimulated Emission of Radiation*]:

The electromagnetic waves experienced in daily life, ranging from the sun, stars, incandescent and fluorescent Lamps are emitted from atoms or molecules **spontaneously**.

- Ordinary natural and artificial, light is released by energy changes on the atomic and molecular level that occur without any outside interference.
- A second type of light exists, when an atom or molecule retains its excess energy until stimulated or induced to emit the energy in the form of light.

Lasers are designed to produce and amplify this stimulated form of light into intense and focused beams. Compared to conventional sources of ordinary light, the light from a laser is quite intense, monochromatic, and emitted in a unidirectional beam limited by diffraction.

The special nature of laser light has made laser technology a vital tool in nearly every aspect of everyday life including communications, entertainment, manufacturing, and medicine.

Stimulated or Induced Absorption:

When an atom absorbs photons of energy $h\nu = E_2 - E_1$ incident on it, in its ground state, the atom reaches in one of its allowed states ***excited state.***

When an atom induced by the photon (energy packet and do transition to one of its allowed states (excited state) is called Stimulated or induced absorption.

The life time of an atom in an excited state is 10^{-8} seconds as shown in Figure 26.10.

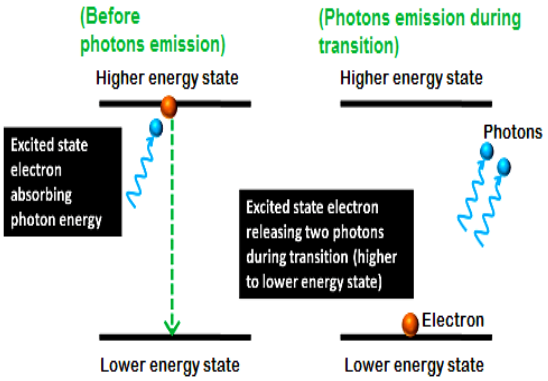


Figure: 26.10. Stimulated Emission

Spontaneous Emission:

The process of photon emission by an excited atom with no external influences is called spontaneous emission.

In excited state the life time of an atom is too short therefore, a probability that the atom in the excited state E_2 will go back to the lower state E_1 by spontaneously emitting the absorbed radiation ($E=h\nu$) is very high. The emitted radiation is incoherent to the induced energy photon as shown in figure 26.11.

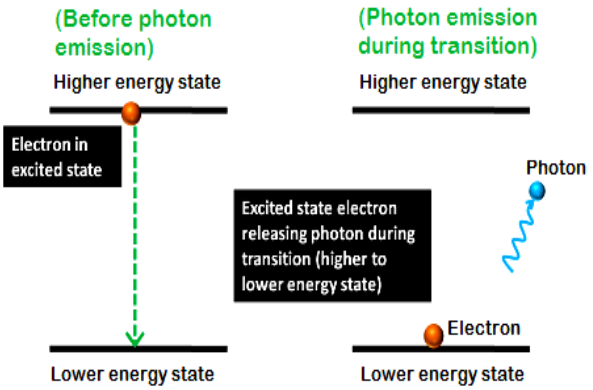


Figure: 26.11. Spontaneous Emission

Stimulated Emission:

In this process, if the atom in an excited state E_2, the action of external radiation with a frequency $h\nu = E_2 - E_1$, forced (induced, stimulated) a transition to the ground state with the emission of one photon with the same energy.

Radiation, which occurs as a result of external exposure, is called induced or stimulated. In the stimulated emission two photons are involved: the primary photon, causing the emission of radiation by an excited atom and the secondary photon emitted by the atom.

These two identical photons will be exactly in phase and coherent. If a cascade of stimulated or induced emission occurs, the number of identical photons increases, the ***light amplification in*** the acronym laser. The beam is *coherent* because all photons are in phase, the beam is *monochromatic* because the photons all have the same *wavelength*, and the beam is parallel because the photons all move in same *direction* as shown in figure 26.12.

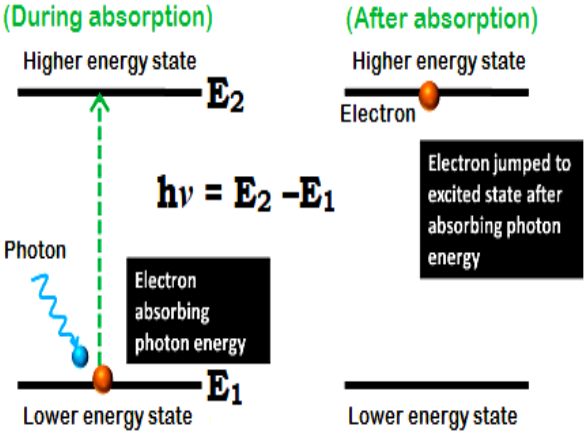


Figure: 26.12. Stimulated Emission

Metastable States and Population Inversion:

Population inversion is a key to produce laser light. The process of achieving a greater population in a higher energy state compared to a lower energy state. The population inversion is required for laser operation.

Population inversion cannot be achieved in a two-energy level system, at normal conditions; the number of electrons in the lower energy state (E_1) is always greater as compared to the number of electrons in the higher energy state (E_2).

DO YOU KNOW?

In addition to induced absorption and spontaneous emission, a third interaction between an atom and a photon (energy) was first proposed by Albert Einstein in 1917 called stimulated emission.

Metastable States:

Consider a system consisting of three energy levels E_1, E_2, E_3. We assume that the energy level of E_1 is less than E_2 and E_3, the energy level of E_2 is greater than E_1 and less than E_3, and the energy level of E_3 is greater than E_1 and E_2. The energy level E_2 is sometimes referred to as **Metastable state** having a life time of 10^{-3} seconds. When we supply light energy which is equal to the energy difference of E_3 and E_1, the electrons in the lower energy state (E_1) gains sufficient energy and jumps into the higher energy state (E_3). This process of supplying energy is called **optical pumping** as shown in figure 26.13.

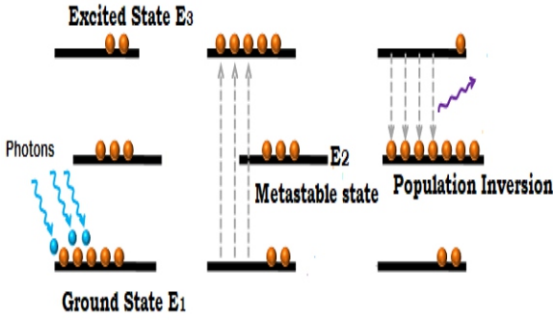


Figure: 26.13. Optical Pumping

We also use other methods to excite ground state electrons such as electric discharge and chemical reactions. The flow of electrons from E_1 to E_3 is called pump transition. Because of the shorter lifetime, only a small number of electrons accumulate in the energy state E_3. The electrons undergo a radiation less transition from E_3 to E_2. In the Metastable state E_2 the electrons will remain there for longer period because of its longer lifetime. Hence the number of electrons coming in the state E_2 is greater than the number of electrons leaving out. As result, a large number of electrons accumulate in Metastable state. After the population inversion is achieved the Metastable state E_2 is exposed to beam of photons which causes induced (stimulated emission). The whole process for population inversion is shown in figure 26.14 for three level systems.

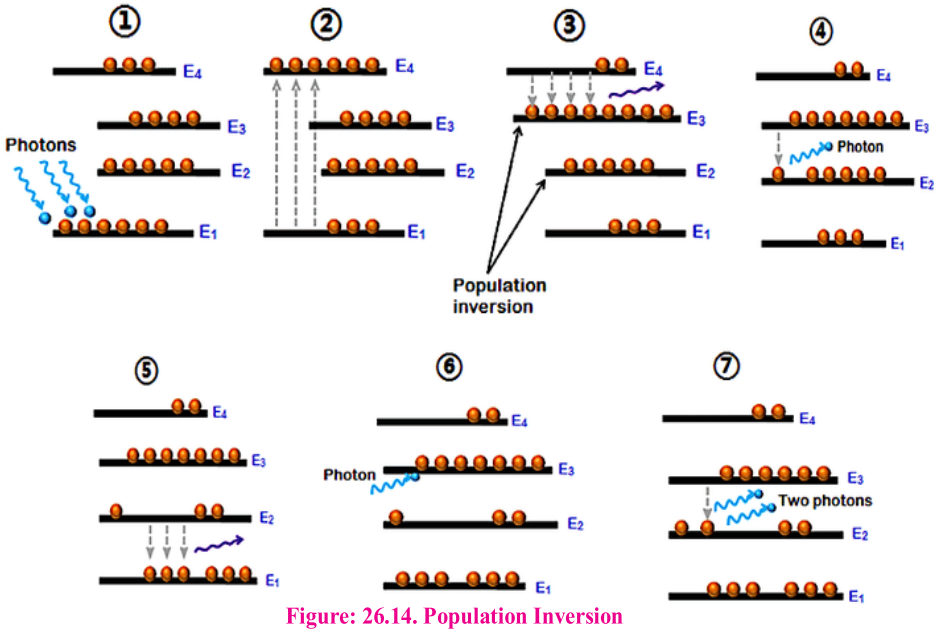


Figure: 26.14. Population Inversion

26.4.2 He – Ne Gas Laser:

The helium-neon laser was the first continuous wave (CW) laser ever constructed. It was built in 1961 by **Ali Javan**, **Bennett**, and **Herriott** at Bell Telephone Laboratories. He-Ne lasers are commonly used in school laboratories and in older barcode readers. The schematic diagram of He-Ne laser is shown in figure 26.15.

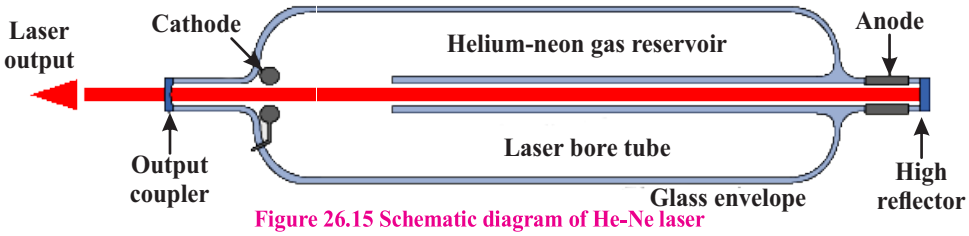


Figure 26.15 Schematic diagram of He-Ne laser

Helium-neon laser construction:

The helium-neon laser consists of three essential components:

Pump Source (high voltage power supply):

The pump energy of the laser is provided by an electrical discharge of several hundred Volts between an anode and cathode at each end of the glass tube. A current of 5 to 100 mA is typical for laser operation.

Gain Medium (laser glass tube or discharge glass tube):

Figure 26.15 shows a gas discharge tube contains a low-pressure mixture of helium-neon in a ratio between 5:1 and 20:1 bound in a glass tube. The partial pressure of helium is 1 mbar whereas that of neon is 0.1 mbar.

Resonating Cavity:

The glass tube (containing a mixture of helium and neon gas) is placed between two parallel mirrors. These two mirrors are silvered or optically coated. Each mirror is silvered differently. The left side mirror is partially silvered and is known as output coupler whereas the right side mirror is fully silvered and is known as the high reflector or fully reflecting mirror. The fully silvered mirror will completely reflect the light whereas the partially silvered mirror will reflect most part of the light but allows some part of the light to produce the laser beam.

DO YOU KNOW?

The three important components of any laser device are:

1. Active medium (medium in which the population inversion can be achieved)
2. Pumping source (usually light or electric current to achieve population inversion)
3. 3. Optical resonator (a feedback system, which consists of an active medium kept in between a 100% mirror and a partial mirror)

Laser Operation:

This electrical excitation raises the helium atom into a meta-stable state with energy 20.61 eV above the ground state. Neon has a meta-stable state with energy 20.66 eV above its ground state. Collision of the excited helium atoms with the ground-state neon atoms results in transfer of energy to the neon atoms, exciting neon electrons. The difference between the energy states of the two atoms is in the order of 0.05eV, which is supplied by kinetic energy. The number of neon atoms in the excited states builds up as further collisions between helium and neon atoms occur, causing a population inversion. Spontaneous and stimulated emission results in emission of 632.82 nm wavelength light.

26.4.3 Properties and Uses of Lasers:

Properties of Laser:

The properties of laser light are very strange as compared to conventional light. The factors that make laser light prominent are: Monochromatic, Highly Intense or Brightness Coherence and Directionality

1. Monochromatic:

A single color or wavelength light is called monochromatic. The light emitted from ordinary light sources has different wavelengths, or colors. However, laser light has a single wavelength or color.

2. Highly Intense or Brightness:

The characteristics of coherence and directionality of the laser makes laser light highly intense as compared to conventional light. It is the power emitted per unit surface area per unit solid angle. A one milli watt He-Ne laser is more intense than the sun intensity.

3. Coherence:

Two or more waves of same frequency are said to be coherent in nature if they have constant phase difference. A predictable correlation of the amplitude and phase at any one point with another point is called coherence. In laser light the property of coherence occurs between any two or more light waves. Where as in conventional light, the property of coherence exhibits between a source and its virtual source.

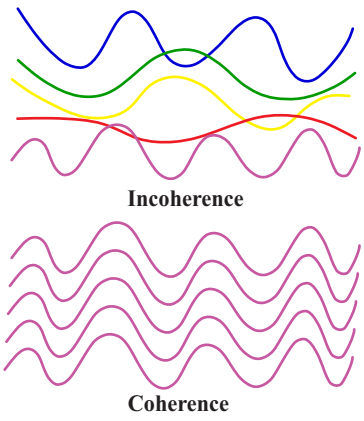


Figure 26.16:
Incoherence and Coherence

4. Directionality or Divergence:

The light ray coming from laser light travels in a single direction. However, an ordinary light source travels in all directions. For example, on travelling a distance of 1 Km the torch light spreads 1 km distance. But the laser light spreads only few centimeters distance even it travels lacks kilometer distance.

Uses of Lasers:

Lasers are widely used in many fields some of them are:

1. Tools:

- Cutting tools are typically precise, and also simple to automate. They do not require sharpening, unlike that of knives.
- Robot-guided lasers are widely used for the cutting of pieces of cloth.
- Laser tools are used in corneal surgery, restoring a detached retina of eye, and removing kidney stones.

2. Communication:

- Lasers are used in Barcode scanners to convert a printed barcode into a number.
- A semiconductor laser beam helps to convert the printed pattern of data into numbers in a CD/DVD.
- In Photonics, lasers are used in fibre optic cables.

3. Defence:

- In defence field, the military employs laser-guided guns and missiles.

- Laser range-finders use high-resolution scanning to find the distance and speed from an object that is located beyond the point-blank range.

4. Medicine:

- In medicine, the laser beam is used as part of phototherapy for many procedures.
- The development of laser technology in recent decades has enabled the creation of a new field of medicine laser surgery.
- Lasers have uses in dermatology, ophthalmology, urology, rheumatology and dentistry.

5. Holography:

- Holography is a true three-dimensional image recording on film by lasers. Holograms are used for amusement, decoration on novelty items and magazine covers.

26.4.3 Safe Handling of Lasers:

Lasers come in various shapes and forms. They have many uses in teaching, research, manufacturing, medicine, dentistry, communications, and shop checkouts and most commonly at work in the office. In fact, some applications may be so well engineered that users are not even aware that the equipment contains a laser.

Laser beam hazards:

The health effects that could occur due to exposure to a laser beam are damage to the skin and eyes. Skin effects include erythema, elastosis (photoageing), and immediate pigment darkening (tanning), burns and skin cancer. Eye effects include photokeratitis, photoconjunctivitis, cataracts, photoretinal damage and burns. The nature of safety measures which are generally taken to avoid the hazards or damages can caused by the lasers are adopted and procedure according to the use and intensity of the laser. Following are some general safety protocols which have to be taken into consideration while working with lasers.

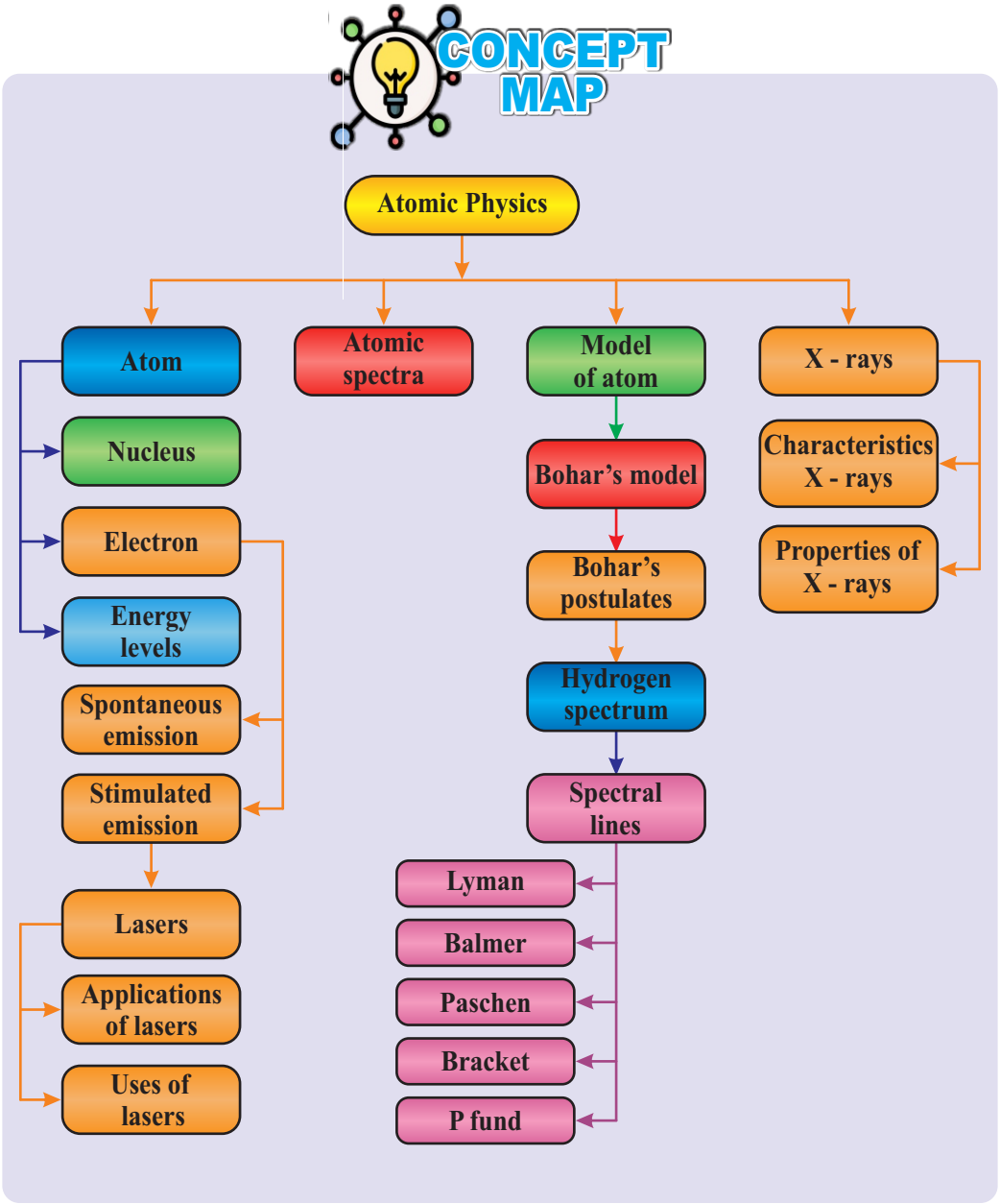
- Only **trained, authorized personnel** may operate lasers, Authorization is received from the authorized laser user and the Laser Safety Officer.
- **NEVER** put yourself into any position where your eyes approach the axis of a laser beam (even with eye protection on).
- **Keep beam paths below or above** standing or sitting eye level. Do not direct them towards other people.
- **Do not damage laser protective housings**, or malfunctioned the interlocks on these housings.
- **Eliminate all reflective material** from the vicinity of the beam paths.
- **Never use viewing instruments** to look directly into a laser beam. If this is necessary, install an appropriate filter into the optical element assembly.
- **Keep ambient light levels as high** as operations will permit.
- **Do not work alone** when performing high power laser operations.

Self-Assessment Questions:

1. How bright is a laser beam when viewed from the side?
2. How do you focus regular light to make it a laser beam?
3. Why do we call laser as a non-material knife?



- ✓ Atomic spectra deal with the spectrum of frequencies of electromagnetic radiation emitted or absorbed by an electron during transitions between different levels of energy within an atom.
- ✓ An optical spectrum is obtained, when light passes through a prism or a diffraction-grating.
- ✓ A line of particular frequency or wavelength emitted or absorbed by atoms is called spectral line.
- ✓ The spectra for each element are unique because each element contains differing numbers of electrons and thus different energy levels.
- ✓ Niels Bohr explained the atom of hydrogen. Hydrogen atom consist of proton and electron revolving around the nucleus
- ✓ According to Bohr's the radii of Hydrogen atom have the quantized values $r = \frac{n^2\varepsilon_0 h^2}{\pi m e^2}$ or $r = r_0 n^2$ where $r_0 = 5.29 \times 10^{-11} m = 0.53 A^0$ is called Bohr's first orbit.
- ✓ X-Rays are powerful electromagnetic rays of higher energy and can pass through most objects, including the body.
- ✓ X-rays are produced when fast electrons, or cathode rays, strike a target, such as the walls or anode of a low-pressure discharge tube.
- ✓ The light from a laser is quite intense, monochromatic, and one directional.
- ✓ The process of photon emission by an excited atom with no external influences is called spontaneous emission.
- ✓ Radiation which occurs as a result of external exposure is called induced or stimulated emission.
- ✓ Metastable state is a particular excited state of an atom, nucleus, or other system that has a longer lifetime than the ordinary excited states.
- ✓ The process of making the population of atoms in the higher energy state (Excited state) more than that of lower energy state (ground state) is known as population inversion.
- ✓ A process in which light is used to raise (pump) electrons from a lower energy level in an atom or molecule to a higher energy level is called optical pumping.
- ✓ The helium-neon laser was the first continuous wave (CW) laser ever constructed.
- ✓ Laser light is Monochromatic, Highly Intense, Coherence and single directional.
- ✓ Lasers are widely used in tools, communication, defence, medicine and holography.
- ✓ The health effects that could occur due to exposure to a laser beam are damage to the skin and eyes.
- ✓ The nature of safety measures which are generally taken to avoid the hazards or damages can caused by the lasers are adopted and procedure according to the use and intensity of the laser.


CONCEPT MAP
Atomic Physics
Atom
Nucleus
Electron
Energy levels
Spontaneous emission
Stimulated emission
Lasers
Applications of lasers
Uses of lasers
Atomic spectra
Model of atom
Bohar's model
Bohar's postulates
Hydrogen spectrum
Spectral lines
Lyman
Balmer
Paschen
Bracket
P fund
X - rays
Characteristics X - rays
Properties of X - rays

EXERCISE

Section (A): Multiple Choice Questions (MCQs)
Choose the correct answer:

1. Atomic spectra are also known as:
   (a) Discrete spectra (b) Line spectra
   (c) Emission spectrum (d) Continuous spectra
2. The ratio of kinetic energy to total energy for an electron in a Bohr orbit is:
   (a) 1:1 (b) 2:3 (c) 1:2 (d) 2:1
3. The Bohr radius increases as the principal quantum number:
   (a) increases (b) decreases
   (c) remains constant (d) oscillates
4. Laser beams consist of:
   (a) Highly coherent electrons (b) Highly coherent photons
   (c) Highly coherent phonons (d) Highly coherent neutrons
5. The ruby laser is an example of:
   (a) Optical pumping (b) Electrical pumping
   (c) Chemical pumping (d) Thermal pumping
6. Laser action requires a medium with at least:
   (a) Three energy levels (b) Four energy levels
   (c) Two energy levels (d) Five energy levels
7. Population inversion occurs in:
   (a) Active medium (b) Passive medium
   (c) Gaseous medium (d) Vapour medium
8. X-rays transfer _____ to metals.
   (a) Energy (b) Force (c) Pressure (d) Momentum
9. X-rays are deflected by:
   (a) Magnetic fields (b) Electric fields
   (c) Gravitational fields (d) No fields
10. Doubling the voltage of an X-ray tube:
    (a) Halves the intensity (b) Keeps the intensity unchanged
    (c) Doubles the intensity (d) Quadruples the intensity

Section (B): CRQs (Short Answered Questions):

1. Why do different elements have different spectra?
2. In the Bohr model, how many times larger is the radius of the fifth Bohr orbit compared to that of the first Bohr orbit?
3. What is the difference between X-Rays and Gamma Rays?
4. Sate the properties of X-Rays, which makes it possible to detect cracks in bones.

5. What is the energy of a photon that, when absorbed by a hydrogen atom, could cause an electronic transition from (a) the n =3 state to the n = 5 state and (b) the n = 5 state to the n = 8 state?
6. What are the (a) wavelength range and (b) frequency range of the Lyman series and the Balmer series?
7. Distinguish between spontaneous and stimulated emission.
8. Explain why population inversion is necessary in a laser?
9. In an optically pumped laser, the light that causes optical pumping is always shorter in wavelength than the laser beam. Explain
10. A hydrogen atom is in its first excited state (n = 2). Calculate (a) the kinetic energy of the electron, (b) the potential energy of the system, and (c) the total energy of the system.

Section (C): ERQs (Long Answered Questions):

1. What are the postulates of Bohr's Model of hydrogen atom? Discuss the importance of this model to explain various series of line spectra in hydrogen atom. Do any of the assumptions of the Bohr's theory of hydrogen atom contradict with the classical Physics? Derive the expression for total energy of electron in nth Bohr orbit and show that- $E_n \propto 1/n^2$.
2. How X-rays are produced? State the purpose of cooling fins in the X-ray tube.
3. Explain why X-rays are appropriate in study of crystalline structure material? Write some main properties of X-rays.
4. What is Laser? Write the characteristics of Laser light. Can a two-level system be used for the production of Laser? Why?
5. What is pumping? What are the different methods of pumping? Explain optical pumping.
6. What is the principle of Laser? Write the construction and working of Helium-neon laser.
7. Give some important properties of lasers. Also write the uses of lasers in the field of medicine, defense and communication.
8. What is wave number? Derive the expression $\frac{1}{\lambda} = R_H\left(\frac{1}{P^2} - \frac{1}{n^2}\right)$.
9. Derive the expression for Bohr's radius and develop a general relation for radii of quantized orbits of hydrogen atom.

Section (D): **Numerical:**

1. Calculate the energy of an electron in the n = 2 orbit of a hydrogen atom according to the Bohr model. **(Ans. $-5.447 \times 10^{-19}\ J$)**
2. Calculate the speed of the electron if it orbits in(a) the smallest allowed orbit and (b) the second smallest orbit? (c)If the electron moves to larger orbits, does its speed increase, decrease, or stay the same? **(Ans. (a) $2.19 \times 10^{6} m/s$ (b) $1.09 \times 10^{6} m/s$ (c) the speed of the electron will decrease)**
3. What are the (a) energy, (b) magnitude of the momentum, and (c) wavelength of the photon emitted when a hydrogen atom undergoes a transition from a state with n = 3 to a state with n = 1? **(Ans. (a) 12.1 eV (b) 6.45 X 10^{-27} N.s (c) 102 nm)**
4. What is the energy of the photon emitted by hydrogen atom when the hydrogen atom changes directly from the n = 5 state to the n = 2 state? **(Ans. 2.85 eV)**
5. How much work must be done to pull apart the electron and the proton that make up the hydrogen atom if the atom is initially in (a) its ground state and (b) the state with n =3?
   (Ans. (a) 13.6 eV (b) 1.51 eV)
6. (a) What is the wavelength of light for the least energetic photon emitted in the Balmer series of the hydrogen atom spectrum lines?
   (b) What is the wavelength of the series limit? **((a) 659 nm (b) 366 nm)**
7. A laser emits light with a wavelength of 632.8 nm and has a power output of 55 mW. Calculate the energy of one photon emitted by this laser. **(Ans. 3.14 X 10^{-19} J)**
8. Calculate the wavelength of X-rays if the energy of one photon emitted by the X-ray machine is 1.9878×10−15 joules. **(Ans. 0.1 nm)**

Unit 27 Nuclear Physics

Teaching Periods 17 **Weightage %** 11



CERN (European Organization for Nuclear Research) is a research organization that operates the largest particle physics laboratory in the world, home to the Large Hadron Collider (LHC). Pakistan has made significant contributions to CERN's LHC project, including designing and manufacturing over 300 superconducting magnets, developing detector components like muon chambers and silicon trackers for the CMS experiment, participating in data analysis and simulations, and establishing a National Grid Computing Center to support LHC research. These contributions showcase Pakistan's expertise in advanced physics research and its collaboration with the global scientific community at CERN.

In this unit student should be able to:

- Recall the composition of atomic nuclei
- Describe isotopes in detail
- Explain that an element can exist in various isotopic forms each with a different number of neutrons.
- Explain the use of mass spectrograph to demonstrate the existence of isotopes and to measure their relative abundance
- Explain the process of radioactive decay
- State law of radioactive decay
- Identify the spontaneous and random nature of nuclear decay.
- Define the terms activity and decay constant and recall and
- Solve problems using $A = \lambda N$
- Infer and sketch the exponential nature of radioactive decay
- Describe the term half life and solve problems using the equation $\lambda = \frac{0.693}{T_{1/2}}$
- Define the terms unified mass scale, mass defect and calculate binding energy using Einstein's equation Illustrate graphically the variation of binding energy per nucleon with the mass number.
- Determine the release of energy from different nuclear reactions.
- Explain that atomic number and mass number conserve in nuclear reactions.
- Describe energy and mass conservation in simple reactions and in radioactive decay.
- Describe the phenomena of nuclear fission and fusion.
- Describe the fission chain reaction
- Describe the function of various components of a nuclear reactor.
- Explain the basic principle of nuclear reactor.
- Describe how the conditions in the interiors of the Sun and other stars allow nuclear fusion to take place and hence, how nuclear fusion is their main energy conversion process.
- Show awareness about nuclear radiation exposure and biological effects of radiation.
- Describe the term dosimetry.
- Describe the use of radiations for medical diagnosis and therapy.
- Explain the importance of limiting exposure to ionizing radiation.
- Describe the examples of the use of radioactive tracers in medical diagnosis, agriculture and industry.

Introduction:

In this chapter, we are going to explore nuclear physics. This field explores the properties and behavior of atomic nuclei, including radioactivity, nuclear reactions, and nuclear energy. First, we will learn about isotopes, which are like different versions of the same atom also see how scientists use mass spectrometers to find these isotopes. Then, we will learn radioactivity, understanding why some atoms emit radiation. Finally, explore the mysteries of nuclear reactions and the cool processes that make nuclei change, even discover how stars and the sun make energy, explore how we can create affordable nuclear energy with reactors, and see how nuclear radiation is used in our everyday lives.

DO YOU KNOW?

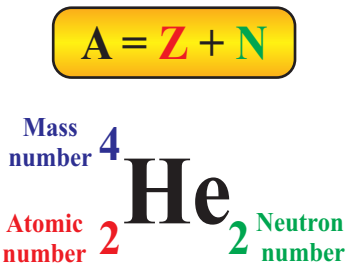
Henri Becquerel discovered uranium's invisible energy, sparking curiosity. Rutherford's bold experiment revealed the dense nucleus at the atom's center, like a strong bowling pin, holding most of the atom's weight and binding it together with the powerful nuclear force

27.1 Atomic Nucleus and Isotope:

The atom is the building block of matter; an atom is composed of a positively charged nucleus, with a cloud of negatively charged electrons surrounding it, bound together by electrostatic force. and atomic nucleus is the small, dense region consisting of protons and neutrons at the center of an atom. Almost all of the mass of an atom is located in the nucleus, with a very small contribution from the electron cloud.

All properties of a nucleus are determined by the number of protons and neutrons it has. A specific combination of protons and neutrons is called a nuclide and is a unique nucleus. The following notation is used to represent a particular nuclide.

$$A = Z + N \quad \text{.....} \quad (27.1)$$



$$^{A}_{Z}X \quad \text{.....} \quad (27.1)$$

Where X shows any nuclide, Z is atomic number and A is mass number. Table 1 lists these quantities and the symbols commonly used to represent them.

For example, a nuclide of aluminum has a mass number (A) of 27 and an atomic number (Z) of 13. Therefore, it has 13 protons and 14 neutrons (27 − 13 = 14) as shown in figure.27.1 and equation

$$A = Z + N \quad \text{.....} \quad (27.2)$$

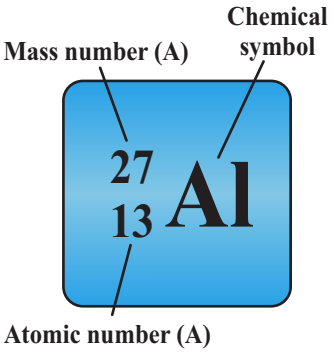


Figure 27.1
Element and its nuclides numbers

Isotopes:

Two or more forms of the same element that contain equal numbers of protons but different numbers of neutrons in their nuclei, and hence they differ in relative atomic mass but not in chemical properties.

Isotopes differ in relative atomic mass but not in chemical properties. For example, carbon has an atomic number of 6. This means that all carbon atoms have 6 protons.
However, carbon has three naturally occurring isotopes: carbon-12, carbon-13, and carbon-14. Carbon-12 has 6 neutrons, carbon-13 has 7 neutrons, and carbon-14 has 8 neutrons.

Isotopes are of two types:

- ***Stable isotopes**:* They do not show any radioactivity and are stable. For example, ${}_7N^{14}$ and ${}_7N^{15}$.
- ***Radioactive isotopes:*** They are unstable and show radioactivity. Some of the radioactive isotopes are ${}_6C^{14}$ for carbon, and ${}_{19}K^{40}$ for potassium.

Isotopic variations: different numbers of neutrons:

An element can exist in various isotopic forms, each with a different number of neutrons. Elements with low atomic numbers tend to have fewer isotopes, while elements with high atomic numbers tend to have more isotopes.
Every element on the periodic table naturally occurs in various isotopic forms, although some may be more prevalent than others. For instance:

- **Oxygen:** About 99.76% of oxygen in nature is Oxygen-16 (with 8 neutrons). However, Oxygen-17 and Oxygen-18 are also found in much smaller quantities.
- **Uranium:** The naturally occurring isotopes of uranium are Uranium-238 (99.3%) and Uranium-235 (0.7%). Uranium has 35 known isotopes. The most stable isotope of uranium is uranium-238, which has 146 neutrons.
- **Hydrogen:** Hydrogen has only three isotopes: protium, deuterium, and tritium. Protium has no neutrons, deuterium has one neutron, and tritium has two neutrons.

DO YOU KNOW?

The Isotopes have same Physical Properties, Despite having different masses like the isotopes of hydrogen have the same physical properties, such as:

- ✓ Boiling point
- ✓ Melting point:
- ✓ Density
- ✓ Solubility:
- ✓ Color: Colorless
- ✓ Odor: Odorless
- ✓ Taste: Tasteless

Self-Assessment Questions:

1. How many protons are there in the nucleus ${}^{197}_{79}Au$?
2. How do isotopes of an element differ from each other?
3. Give an example of two isotopes of the same element.

27.1.4 Mass Spectrograph:

A mass spectrograph is a device that can be used to separate isotopes of an element based on their mass and it works by accelerating charged particles through a magnetic field. The particles are then deflected by the magnetic field, and the amount of deflection is determined by the mass of the particle.

Bainbridge mass spectrograph is a type of mass spectrometer that uses a combination of electric and magnetic fields to separate Ions according to their charge-to-mass ratio. It is named after its inventor, Kenneth T. Bainbridge, who developed it in 1933.

Principle:

The fundamental principle behind the Bainbridge Mass Spectrograph is the application of magnetic and electric fields to separate and measure the masses of charged particles, typically Ions. This separation is based on the principles of magnetic deflection and kinetic energy.

The **Bainbridge mass spectrograph** consists of three main components: an ion source, a velocity selector, and a magnetic analyzer.

- The **ion source** is where the Ions are produced. The Ions are typically produced by bombarding a sample with electrons, which knocks electrons out of the atoms, creating positively charged Ions.
- The **velocity selector** is used to select Ions of a particular velocity. The Ions are passed through a region of electric and magnetic fields, which are adjusted so that only Ions of a certain velocity can pass through.
- The **magnetic analyzer** is used to separate the Ions according to their charge-to-mass ratio. The Ions are passed through a region of magnetic field, and the radius of their path is determined by their charge-to-mass ratio.

The Ions that are separated by the magnetic analyzer are then detected by a detector, such as a photographic plate or a computer.

A schematic diagram showing the construction of Bainbridge's mass spectrograph is shown in Figure 27.2.A beam of positive Ions produced in a discharge tube is collimated into a narrow beam by the two slits S_1 and S_2. After emerging from the slit, the positive Ions enter a velocity selector.

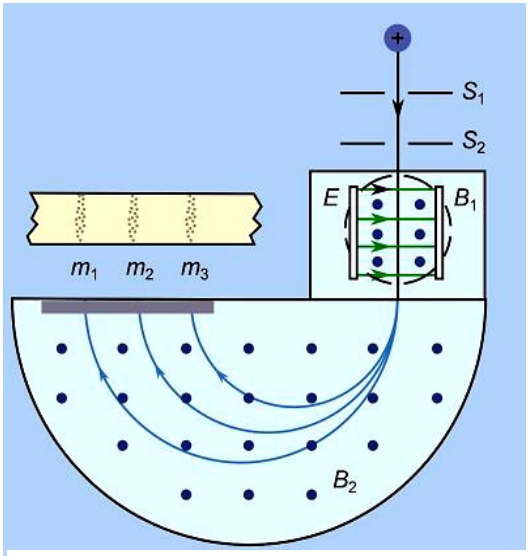


Figure 27.2 Bainbridge's mass spectrograph

The velocity selector consists of two plates E and B_1, between which a steady electric field is maintained in a direction at right angles to the ion beam. The electric field and the magnetic field of the velocity selector are so adjusted that the deflection produced by one is exactly equal and opposite to the deflection produced by the other so that there is no net deflection for Ions having a particular

$$Xe = B_1 ev \ldots 27.3$$

$$v = \frac{X}{B_1} \ldots\ldots\ldots\ldots\ldots 27.4$$

The Ions were accelerated to a known kinetic energy, ensuring that all Ions of different masses had the same kinetic energy. Positive ions entering the evacuated chamber are subjected to a perpendicular electromagnetic field of intensity B_2, causing them to follow

curved paths. The curvature depends on their charge-to-mass ratio (e/m), with heavier ions curving less than lighter ones. This method is very accurate due to the linear mass scale. When a charged particle with mass m and charge e is accelerated through a potential difference V, it gains a velocity v, given by

$$\frac{1}{2}mv^2 = eV \quad or \quad v = \sqrt{\frac{2eV}{m}} \ldots\ldots\ldots\ldots 27.5$$

When this charged particle moving with a velocity v enters a magnetic field B in a direction perpendicular to the field, the force acting on it due to the field is Bev acting in a direction at right angles to the direction of motion of the charged particle and that of the magnetic field. The particle, therefore, moves along a circular path of radius r given by

$$Bev = \frac{mv^2}{r} \ldots\ldots\ldots 27.6$$

Rearrange eq:27.6 for mass m and putting value of v from equation 27.5 $v = \sqrt{\frac{2eV}{m}}$),

$$m = \frac{Ber}{v} = \frac{Ber}{\sqrt{\frac{2eV}{m}}} \ldots\ldots\ldots 27.7$$

Squaring both sides and after re-arranging,

$$m = \left(\frac{er^2}{2V}\right)B^2 \ldots\ldots\ldots\ldots 27.8$$

The mass of each ion reaching the detector depends on the value of B^2. By changing B and keeping other variables constant, ions of different masses can be directed into the detector. This allows us to create a graph of detector readings versus B^2, identifying the masses and quantities of the ions present. Multiple peaks in the mass spectrum indicate different isotopes of the same element.

For instance, chlorine's mass spectrum shows two peaks corresponding to chlorine-35 and chlorine-37, demonstrating the presence of isotopes. The height of each peak reveals the relative abundance of each isotope, with more abundant isotopes producing taller peaks and less abundant ones producing shorter peaks. Comparing peak intensities helps determine the relative abundance of the isotopes.

27.2 Radioactive Decay:

Radioactivity is the process by which an unstable atomic nucleus loses energy by radiation. A material containing unstable nuclei is considered radioactive.

There are two types of radioactivity:

1. **Natural Radioactivity:** Natural radioactivity is the radioactivity that occurs naturally in the environment. It is caused by the decay of unstable isotopes of elements that are found in the Earth's crust, such as uranium, thorium, and potassium-40.

DO YOU KNOW?

Radioactivity was first discovered by Henri Becquerel back in 1896. He observed that a piece of uranium salt, wrapped in paper, emitted certain strong rays that could affect a photographic plate.

2. **Induced Radioactivity:** Induced radioactivity is the radioactivity that is created by bombarding a stable material with radiation, such as neutrons or protons. This can be done in a nuclear reactor or particle accelerator.

The main difference between natural and induced radioactivity is the source of the radiation. In natural radioactivity, the radiation comes from the unstable nuclei of the atoms themselves. In induced radioactivity, the radiation comes from outside the atoms, and is used to destabilize their nuclei.

Henri Becquerel discovery of radioactivity was that the rays were emitted without any external stimulation or excitation. In other words, uranium salt was naturally radioactive, meaning it produced these penetrating radiations all on its own, without any external forces influencing it. Further research by scientists like Madame Curie, Pierre Curie, and Rutherford confirmed that this phenomenon wasn't unique to uranium; it could also be observed in heavy elements like polonium, radium, and thorium.

27.2.1 Radioactive Decay Process:

The natural radioactivity observed was specific to heavy elements with atomic weights greater than about 206. Of the 2,500 identified types of atoms, around 90% are radioactive, meaning they change into other types of atoms over time. These unstable nuclides release energy and particles in much different way of decay types to become stable nuclides.

Alpha Decay:

All atomic nuclei with a proton number (Z) greater than 83 and a mass number (A) greater than 209 undergo spontaneous transformation into lighter nuclei by emitting one or more alpha particles, which are equivalent to helium-4 nuclei.

$$^{A}_{Z}X \rightarrow ^{A-4}_{Z-2}Y + ^{4}_{2}He \ldots\ldots\ldots\ldots\ldots 27.9$$

where X is parent nuclei,Y is daughter nuclei and $^{4}_{2}He$ is alpha particle. Examples of alpha decays are:

$$^{238}_{92}U \rightarrow ^{234}_{90}Th + ^{4}_{2}He$$

In this example, an isotope U $^{-238}$ converts into Th $^{-90}$ with emission of alpha particle. Alpha decay is possible whenever the mass of the original neutral atom is greater than the sum of the masses of the final neutral atom and the neutral helium-4 atom.

Worked Example 27.1

The element radium was discovered by Marie and Pierre Curie in 1898.One of the isotopes of radium, $^{226}_{88}Ra$, decays by alpha emission. What isthe resulting daughter element?

Solution:

Step 1: Write down the known quantities and quantities to be found.

Given: The decay can be written symbolically as follows:

$$^{226}_{88}Ra \rightarrow Y + ^{4}_{2}He$$

Unknown: the daughter element (Y).

The mass numbers and atomic numbers on the two sides of the expression must be the same so that both charge and nucleon number are conserved during the course of this particular decay.

Mass number of Y = 226 − 4 = 222
Atomic number of Y =88 – 2=86

$$^{226}_{88}Ra \rightarrow {}^{222}_{86}Y + {}^{4}_{2}He$$

Step 2: The periodic table shows that the nucleus with an atomic number of 86 is radon,
Result: Thus, the decay process is as follows:

$$^{226}_{88}Ra \rightarrow {}^{222}_{86}Rn + {}^{4}_{2}He$$

In this process, radium (Ra^{-226}) transforms into radon (Ra^{-222}).

Beta Decay:

The beta-decay is a spontaneous process in which mass number of the nucleus remains unchanged, but the atomic number changes by unity ($\Delta Z = \pm 1$). The change in atomic number is due to emission of an electron, emission of positron or by capture of an orbital electron (K−capture). If the daughter nucleus doesn't have the right ratio of neutrons and protons to stay stable, it can change through beta decay to become more stable. There are three types of β−decay processes depending upon their mode of decay: β^-−decay (electron emission), β^+−decay (positron emission) and electron capture.

β^-−Decay (Electron Emission): General equation of electron emission is:

$$^{A}_{Z}X \rightarrow {}_{Z+1}{}^{A}Y + {}^{0}_{-1}e \ldots\ldots\ldots 27.10$$

Whenever, the number of neutron is more than the number of proton then negative beta decay happens. In this process, a neutron is transformed into a proton:

$$^{1}_{0}n \rightarrow {}^{1}_{1}p + {}^{0}_{-1}e \ldots\ldots\ldots 27.11$$

Here is an example:

$$^{14}_{6}C \rightarrow {}^{14}_{7}N + {}^{0}_{-1}e$$

This equation represents the decay of carbon-14 ($_6C^{14}$) into nitrogen-14 ($_7N^{14}$) with the emission of an electron.

β^+−Decay (Positron Emission):

Also known as positive beta decay. General equation of positron emission is:

$$^{A}_{Z}X \rightarrow {}_{Z-1}{}^{A}Y + {}^{0}_{+1}e \ldots\ldots\ldots 27.12$$

Whenever, the number of proton is more than the number of neutron then positive beta decay happens. In this process, a proton is transformed into a neutron:

$$^{1}_{1}p \rightarrow {}^{1}_{0}n + {}^{0}_{+1}e \ldots\ldots\ldots 27.13$$

Here is an example:

$$^{11}_{6}C \rightarrow {}^{11}_{5}B + {}^{0}_{+1}e$$

This equation represents the decay of carbon-11 ($_6C^{11}$) into boron-11 ($_5B^{11}$) with the emission of a positron (e^+).

Electron Capture: In this process an electron from K-shell of the atom is captured by the nucleus to form a new nucleus and a photon is emitted:

$$^{A}_{Z}X + {}^{0}_{-1}e \rightarrow {}_{Z-1}{}^{A}Y + h\nu \ldots\ldots\ldots 27.14$$

In this process, atomic number Z is decreased by one but mass number A remains the same. In this case a proton in the nucleus is converted into a neutron by capture of an orbital electron:

$$ {}_{1}^{1}p + {}_{-1}^{0}e \rightarrow {}_{0}^{1}n \ldots\ldots\ldots\ldots 27.15 $$

An example of electron capture is the decay of Krypton-81 into Bromine-81:

$$ {}_{36}^{81}K + {}_{-1}^{0}e \rightarrow {}_{35}^{81}Br $$

In this reaction, the atomic number drops from 36 to 35, while the mass number stays at 81.

Gamma Decay:

The gamma decay is the spontaneous emission of electromagnetic photon from the nucleus. The emission of alpha or beta particles from the natural radioactive substance may leave the daughter nucleus in one or more excited states. When the nucleus in excited states goes to lower or ground state, then it emits gamma rays. During gamma decay, there is no change in mass number A or atomic number Z. An excited nucleus is denoted by an asterisk (*) after or over its usual symbol. Thus $_{38}S^{87*}$ refers to $_{38}S^{87}$ in an excited state. The general equation for gamma decay is:

$$ {}_{Z}^{A}X^{*} \rightarrow {}_{Z}^{A}X + \gamma \ldots\ldots\ldots 27.16 $$

Here is an example of gamma-decay:

$(_{27}Co^{60*}) \rightarrow {}_{27}A^{60} + \gamma$

This equation represents the decay of an excited state of cobalt-60 ($_{27}Co^{60*}$) into the ground state of cobalt-60 ($_{27}Co^{60}$) with the emission of a gamma (γ) photon.

The summary of the radiations from nucleus is given in Table 27.1.

Table 27.1 Alpha, Beta and Gamma radiation

Particle	Symbols	Composition	Charge	Effect on parent nucleus
alpha	$\alpha\ ({}_{2}^{4}He)$	2 protons, 2 neutrons	+ 2	mass loss; new element produced
beta	$\beta^{-}\ ({}_{-1}^{0}e)$	electron	− 1	no change in mass number; new element produced
	$\beta^{+}\ ({}_{1}^{0}e)$	positron	+ 1	
gamma	γ	Photon	0	energy loss

Self-Assessment Questions:

The isotope $_{26}^{56}Fe$ decays into the isotope $_{27}^{56}Co$.

1. By what process will this decay occur?
2. Write the decay formula for this process.

27.2.2 Law of Radioactive Decay:

The law of radioactive decay states that the rate of decay of a radioactive sample at any instant is directly proportional to the number of atoms present at that instant.

It implies that the larger the quantity of the radioactive material, the faster it will decay.

Let ΔN be the number of atoms disintegrating in a time Δt and N be the number of atoms present at that instant, then the rate of decrease $-\frac{\Delta N}{\Delta t}$ is proportional to N,

$$\frac{\Delta N}{\Delta t} \propto -N \quad , \qquad \frac{\Delta N}{\Delta t} = -\lambda N \ldots\ldots 27.17$$

Where negative sign indicates that there is decrease in N with time and λ is called ***decay constant*** or disintegration constant. It is defined as the ratio of the amount of the substance which disintegrates in a unit time to the amount of substance present. Isotopes with a large value of λ decay rapidly; those with small λ decay slowly.

We can write the above equation in derivative form as $dN/N = -\lambda dt$. Let the number of radioactive atoms initially present be N_0 at t = 0; and at some time t, the number of radioactive atoms are N.

Applying the limits and integrating equation both sides:

Taking anti logarithm on both sides of above equation:

$$\frac{N}{N_0} = e^{-\lambda t}$$

$$N = N_0 e^{-\lambda t} \ldots\ldots\ldots 27.18$$

This equation shows that the number of atoms of a given radioactive substance decreases exponentially with time.

DO YOU KNOW?

The key thing to remember is that this natural radioactivity is entirely spontaneous and not affected by external factors. Artificial radioactivity is also spontaneous in nature.

27.2.3 The Spontaneous and Random Nature of Radioactive Decay:

Radioactive decay is both spontaneous and random.

Spontaneous process:

It is a process which cannot be influenced by environmental factors, Such as:

(i) Temperature
(ii) Pressure
(iii) Chemical conditions

Random process:

It is a process in which the exact time of decay of a nucleus cannot be predicted

The random nature of radioactive decay can be demonstrated by observing the count rate of a Geiger-Muller (GM) tube.

(i) When a GM tube is placed near a radioactive source, the counts are found to be irregular and cannot be predicted.
(ii) Each count represents a decay of an unstable nucleus.
(iii) These fluctuations in count rate on the GM tube provide evidence for the randomness of radioactive decay.

The nucleus has a constant probability, ie. the same chance, of decaying in a given time. Therefore, with large numbers of nuclei, it is possible to statistically predict the behavior of the entire group.

27.2.4 Activity (A):

Sometimes we are more interested in the decay rate A ($A = -\frac{\Delta N}{\Delta t}$) than in N itself. The **activity A** of a radioactive sample is the number of disintegrations (decays) occurring per unit of time.

If a radioactive sample contains N atoms at any time t, then its activity at time t is given as

$$A = -\frac{\Delta N}{\Delta t} \ldots\ldots\ldots 27.19$$

where negative sign shows that activity decreases with time. According to law of radioactive decay

$$\frac{\Delta N}{\Delta t} = -\lambda N$$

Therefore,

$$A = \lambda N$$

Since $N = N_0 e^{-\lambda t}$, so

$$A = \lambda N_0 e^{-\lambda t}$$

Putting $A_o = \lambda N_o$ in above equation, we get

$$A = A_0 e^{-\lambda t} \ldots\ldots\ldots 27.20$$

The equation $A = \lambda N$ *and* $A = A_0 e^{-\lambda t}$ are alternative forms of the law of radioactive decay. The total decay rate A of a sample of one or more radionuclide is called the activity of that sample. The SI unit for activity is the Becquerel (Bq), which corresponds to one disintegration per second. Another common unit is the curie (Ci), where 1 Ci is approximately equal to 3.7×10^{10} Bq.

Worked Example 27.2

A radioactive sample of radium ($^{226}_{88}Ra$) contains 3×10^{16} nuclei. The decay constant λ is $1.4 \times 10^{-11} s^{-1}$. What is the activity A of the sample? Convert answer in curies.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Number of nuclei $N = 3 \times 10^{16}$, $\lambda = 1.4 \times 10^{-11} s^{-1}$

Unknown: A = ?

The activity A of the sample is given by:

$$A = \lambda N$$

Step 2: Putting the values,

$$A = 1.4 \times 10^{-11} \times 3 \times 10^{16} = (1.4 \times 3) \times 10^{-11+16} = 4.5 \times 10^{5}$$

$$A = 4.5 \times 10^{5} \text{Bq}$$

Step 3: In order to convert above value in curies, we will divide it by the factor 3.7×10^{10} *Bq/Ci:*

$$A = \frac{4.5 \times 10^5 Bq}{3.7 \times \frac{10^{10} Bq}{Ci}} = 1.1 \times 10^{-5} Ci$$

Result: **A**=$1.1 \times 10^{-5} Ci$

27.2.6 Exponential Nature of Radioactivity:

Radioactive decay follows an exponential decay pattern. This means that the rate of decay proportionally decreases with the remaining amount of radioactive material and it can be visualized graphically by decay curve.

A decay curve is a plot of the number of radioactive parent nuclei remaining in a sample as a function of half-life. A typical decay curve for a radioactive sample is shown in Figure 27.3. After each half-life, half the remaining parent nuclei have decayed. This is represented in the circles to the right of the decay curve. The blue spheres are the parent nuclei (carbon-14), and the red spheres are daughter nuclei (nitrogen-14). Notice that the total number of nuclei remains constant, while the number of carbon atoms continually decreases over time.

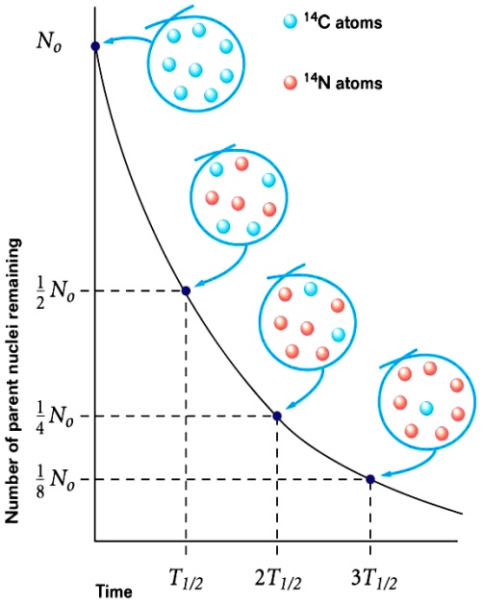


Figure: 27.3
decay curve for a radioactive sample

For example, the initial sample contains 8 carbon-14 atoms. After one half-life, there are 4 carbon-14 atoms and 4 nitrogen-14 atoms. By the next half-life, the number of carbon-14 atoms is reduced to 2, and the process continues. As the number of carbon-14 atoms decreases, the number of nitrogen-14 atoms increases.

The curve gets shallower and shallower as time goes on. This is because there are fewer and fewer radioactive nuclei left to decay, so the rate of decay slows down. This exponential decay is a characteristic feature of radioactive substances and is commonly used in radiocarbon dating and other applications in nuclear physics.

The key features of the exponential nature of radioactive decay include:

- The rate at which radioactive atoms decay is proportional to the quantity of the substance present.
- When plotted on a graph, the decay curve is a smooth, continuous curve that approaches but never reaches zero.
- The curve demonstrates a steeper decline at the beginning (short times) and a gradual decrease as time progresses.
- While we can't predict when a specific atom will decay, we can predict the probability of decay within a certain time frame.

27.2.7 Half-Life (t1/2):

The half-life of a radioactive substance is the time it takes for half of the initial amount of the substance to decay or transform into another element. It's a measure of the stability of an atom and the rate at which it loses its radioactivity.

Decay Constant (λ):

The decay constant, is a measure of the rate at which a radioactive substance decays. It's denoted by the symbol λ and represents the probability of decay per unit time.

Formula:

The half-life formula is:

$t1/2 = \ln(2) / \lambda$

Where:

- t1/2 is the half-life
- ln(2) is the natural logarithm of 2 (approximately 0.693)

λ is the half-life constant (decay constant)

This shows that the half-life is inversely proportional to the decay constant. A higher decay constant means a shorter half-life, and vice versa.

Worked Example 27.3

A 50.0-g sample of carbon is taken from the pelvis bone of a skeleton and is found to have a carbon-14 decay rate of 200.0decays/min. It is known that carbon from a living organism has a decay rate of 15.0 decays/min-g and that C-14 has a half-life of 5730 years = 3.01×10^9 min. Find the age of the skeleton.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Given:$A = 200$ decays/min

Calculate the original activity A_0 from the decay rate and the mass of the sample:

$A_0 = 15$ decay/min-g $\times$ 50g $= 7.5 \times 10^2$ decays/min

$T_{1/2} = 3.01 \times 10^9$min

Unknown: t = ?

Step 2: Age of skeleton can be found by the equation:

$$A = A_0 e^{-\lambda t}$$

Re-arranging,

$$t = \frac{\left(ln\frac{A}{A_0}\right)}{\lambda}$$

Step 3: Find the decay constant from the half-life:

$$\lambda = \frac{0.693}{T_{1/2}} = \frac{0.693}{3.01 \times 10^9} = 2.3 \times 10^{-10} min$$

Step 4: Age will be:

$$t = \frac{\left(ln\frac{200}{7.5\times10^{2}}\right)}{2.3 \times 10^{-10}} = \frac{1.32}{2.3 \times 10^{-10}} = 5.74 \times 10^{9} min$$

Result: $t = 1.09 \times 10^{4}$ years.

Self-Assessment Questions:

1. The half-life of $^{131}_{53}I$ is 8.07 days. Calculate the decay constant for this isotope. What is the activity in Ci for a sample that contains 2.5×10^{10} iodine-131 nuclei?
2. A radioactive sample consists of 5.3×10^{5} nuclei. There is one decay every 4.2 hrs.
   (a) What is the decay constant for the sample?
   (b) What is the half-life for the sample?

27.3 Mass Defect and Binding Energy:

When examining the nucleus of an atom, we find that its mass is actually less than the combined mass of its protons and neutrons. This discrepancy is due to the strong nuclear forces holding the nucleus together. To better understand this phenomenon, we need to learn some important terms.

Mass Defect and Binding Energy or the Hidden Energy within Atoms:

Two fundamental concepts that play a critical role in understanding atomic behavior are Mass Defect and Binding Energy.

Mass Defect:

Mass Defect refers to the difference between the total mass of individual protons, neutrons, and electrons in an atom and the actual mass of the atom.

Binding Energy:

Binding Energy is the energy required to break apart the nucleus of an atom into its constituent protons and neutrons. Exploring the Connection: as shown in figure 27.4

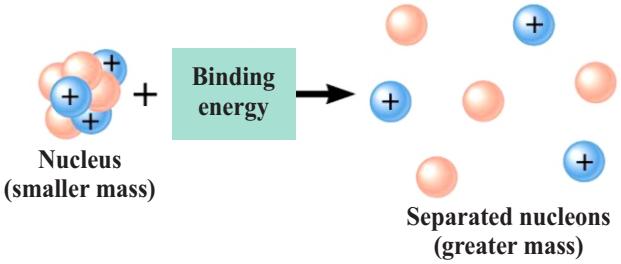


Figure 27.4. mass defect and binding energy of an atom

27.3.1 Unified Mass Scale and Mass Defect:

Unified mass scale:

It is also known as the **Unified Mass Scale** or **Atomic Mass Unit (amu)**, is a unit of measurement used in atomic and nuclear physics to express the mass of atoms and molecules on a scale that is convenient for working with atomic and molecular masses based on the carbon-12 isotope.

The Atomic Mass Unit (a.m.u) is defined as 1/12th of the mass of $_6C^{12}$ atom (C−12). According to Avogadro's Hypothesis, the number of atoms in 12 gm of $_6C^{12}$ is equal to Avogadro number i.e., 6.023×10^{23}.

The mass of one carbon atom $= \frac{12}{6.023\times10^{23}} = 1.992678 \times 10^{-26}$kg.

Therefore 1 a.m.u is defined as:

1a.m.u $= \frac{1}{12} \times 1.992678 \times 10^{-26} = 1.660565 \times 10^{-27}$kg

For examples:

- Mass of a proton $m_p = 1.007276$ a.m.u$= 1.67865 \times 10^{-27}$kg
- Mass of a neutron$m_n = 1.008665$ a.m.u$= 1.67495 \times 10^{-27}$kg
- Hydrogen atom $_1H^1 = 1.00784$ a.m.u
- Chlorine atom = 35.47 a.m.u

Energy equivalent of a.m.u:

According to Einstein mass-energy relation, the energy equivalent to mass m given by: $E = mc^2$ where c is velocity of light.

Let mass m = 1 a.m.u $= 1.680665 \times 10^{-27}$kg, c $= 3 \times 10^8$ m/s. Then energy equivalent to 1 a.m.u is given as:

1 a.m.u $= 1.680666 \times 10^{-27} \times (8 \times 10^8)^2 = 1.4925 \times 10^{-10}$ Joules

1 a.m.u$= \frac{1.4925\times10^{-10}}{1.602\times10^{-13}} = 931.5$ MeV

Where 1MeV $=1.602\times10^{-13}$J. Energy equivalent to mass of electron, proton and neutron are:

$m_e = 0.511$MeV, $m_p = 938.279$MeV and $m_n = 939.573$MeV.

The mass and rest energy of all particles are given in the table 27.2.

Table 27.2 Mass and Rest energy of atomic particles

Particle	m (kg)	m (u)	E_R (MeV)
Proton	1.673×10^{-27}	1.007276	938.3
Neutron	1.675×10^{-27}	1.008665	939.6
Electron	9.109×10^{-31}	1.000549	0.5110

Binding Energy and Mass Defect:

Binding energy is the energy required to separate all the protons and neutrons in a nucleus as shown in figure 27.5. It is a measure of the stability of the nucleus. The mass defect, which is the difference between the mass of the nucleus and the sum of the individual masses of its protons and neutrons, is converted into binding energy according to Einstein's equation $E= mc^2$.

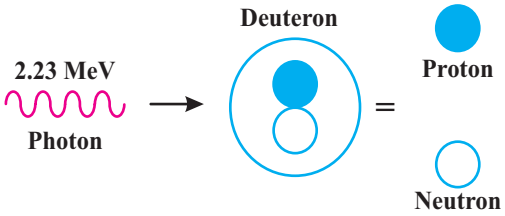


Figure 27.5 Splitting of nucleus

This energy is what holds the nucleus together, overcoming the repulsive forces between the positively charged protons. The mass of the nucleus so formed is less than the sum of the masses of the constituent protons and neutrons. This mass difference is called **mass defect** and is denoted by Δm.

If Z is the number of protons in the nucleus (Atomic number Z) and A is the atomic mass, then the number of neutrons in the nucleus is (A - Z). If m_p is the mass of the proton and m_n is the mass of the neutron, then:

Sum of the masses of the protons and neutrons = $Zm_p + (A - Z)m_n$.

If M_N is the actual mass of the nucleus, then Mass defect Δm is given by,

$$\Delta m = Zm_p + (A - Z)m_n - M_N \ldots\ldots\ldots 27.21$$

If mass defect is Δm then the binding energy of the nucleus according to Einstein mass energy relation is given as:

Binding Energy = $E_B = \Delta mc^2$

$E_B = c^2[Zm_p+(A-Z)m_n-M_N]$

where M_N is the mass of the nucleus, m_p the mass of the proton and m_n the mass of the neutron.

Worked Example 27.5

Calculate the mass defect and binding energy of deutron which contains one proton and one neutron. The mass of deutron nucleus is 2.0136 a.m.u, mass of a proton is 1.0073 a.m.u and mass of a neutron is 1.0087 a.m.u.

Solution:

Step 1: Write down the known quantities and quantities to be found.

mass of proton:1.0073 a.m.u, mass of neutron = 1.0087 a.m.u and mass of deutron = 2.0136 a.m.u.

Unknown: Δm = ?, E_B =?

Step 2: The combined mass is of proton and neutron = [1.0073 + 1.0087] = 2.1060 a.m.u.

Step 3: Therefore, mass defect: Δm = 2.1060 − 2.0136 = 0.0024 a.m.u.

The biding energy is given by:

$$E_B = \Delta m \times 931.5 = 0.0024 = 2.2356 MeV$$

Result: $E_B = 2.2356 MeV$

27.3.2 Binding Energy Per Nucleon:

The binding energy per nucleon ($\overline{E_B}$) is the average energy required to release a nucleon from the nucleus. It is given mathematically as

$$\overline{E_B} = \frac{E_B}{A} \ldots\ldots\ldots 27.22$$

where A is the total number of nucleons in the nucleus.

The binding energy per nucleus$\overline{E_B}$ of a nucleus is more important than total binding energy E_B of the nucleus because $\overline{E_B}$determines the stability of the nucleus. Higher is the value of the binding energy per nucleon, more stable is the nucleus. The nuclei with less binding energy per nucleon are comparatives less stable.

The graph between binding energy per nucleon and mass number of different nuclei is shown in Figure 27.6. Following are the main features of binding energy curve:

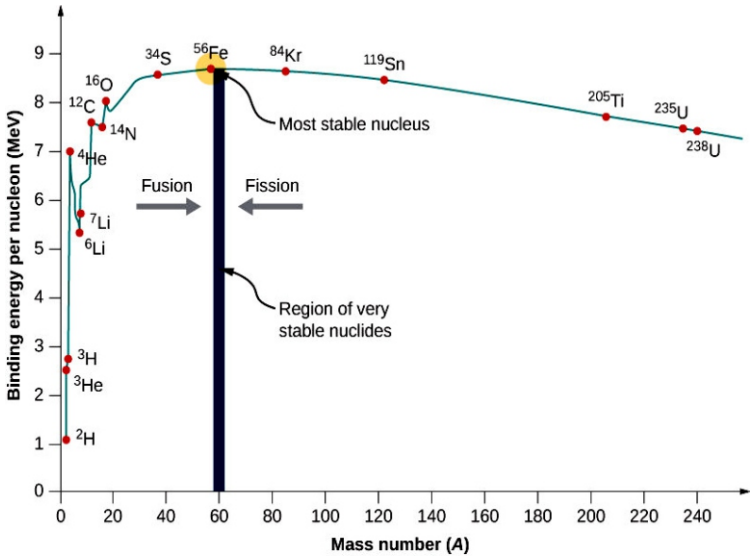


Figure 27.6 Binding energy per nucleon and mass number of different nuclei

- The binding energy per nucleon for light nuclei (i.e. $_{a}H^{2}$) is very small. The binding energy per nucleon increases rapidly for nuclei up to mass number 20.
- The curve possesses peaks corresponding to nuclei $_{2}He^{4}$,$_{4}Be^{8}$, $_{6}C^{12}$, $_{8}O^{16}$ and $_{10}Ne^{20}$.
- The peaks indicate that these nuclei are more stable than other nuclei in their neighborhood. ù After mass number 20 binding energy per nucleon increases gradually.
- The curve has average value of binding energy per nucleon of about 8.6 MeV for a very considerable range of mass number 10 to 120. In this range the curve is more or less flat.
- For mass number A = 6(i.e., $_{26}F^{56}$), the binding energy per nucleon is maximum and is equal to 8.8 MeV.

- The nuclei of intermediate masses (i.e. A=40 to 120) are most stable and a very high amount of energy has to be supplied to liberate each of their nucleons.
- The binding energy per nucleon has a low value for both very light and very heavy nuclei.
- In order to attain higher value of binding energy per nucleon, the lighter nuclei may unite together to form a heavier nucleus (process of nuclear fusion) or a heavier nucleus may split into light nuclei (process of nuclear fission). In both these nuclear processes, the resulting nucleus acquires greater value of binding energy per nucleon and large amount of energy is released.

27.4 Nuclear Reactions:

Any process that involves a change in the nucleus of an atom is called a nuclear reaction. Mathematically,

$$X + x \rightarrow Y + y + Q$$

where X is the target, x are the projectiles, y are the ejectiles and Y is called the residual(product) nucleus.

DO YOU KNOW?

Rutherford produced the first nuclear reaction in a laboratory experiment in 1919. He used 7.7 MeV alpha particles from the decay of ^{210}Po radioactive source on nitrogen and demonstrated the following nuclear changes:
$^{14}_{7}N + ^{4}_{2}He \rightarrow ^{17}_{8}O + ^{1}_{1}H + Q$
The above equation is called Nuclear Reaction.

27.4.1 Energy Released from Nuclear Reactions:

The energy is either absorbed or emitted which is called **Q value** and it is equal to the mass defect. The Q value can also be define as the difference between the rest energies of X and x and the rest energies of Y and y:

$$Q = (m_X + m_x - m_Y - m_y)c^2 \ldots\ldots\ldots\ldots 27.23$$

The value of Q is taken as positive when the energy is released, and its corresponding reaction is called **exothermic.** Similarly, the value of Q is taken as negative when the energy is absorbed, and its corresponding reaction is called **endothermic**.

27.4.2 Conservation of Atomic and Mass Numbers:

Whenever there is a nuclear decay, there is always some physical quantities need to be conserved or remain constant. The following are rules for any nuclear reaction:

1. The total of the atomic numbers (Z) on the left is the same as the total on the right of the given equation because charge must be conserved.
2. The total of the mass numbers (A) on the left is the same as the total on the right of equation because nucleon number must be conserved.

For example, in following nuclear reaction, the total atomic number and mass number remain same on both sides of the equation:

$$^{A}_{Z}X = ^{238}_{92}U \rightarrow ^{234}_{90}Th + ^{4}_{2}He \ldots\ldots\ldots 27.24$$

In above reactions, Z=92 on the left side which is equal to 90+2=92 on the right side. Similarly, A=238 on the left side which is equal to 234+4=238 on the right side.

27.4.3 Conservation of Mass and Energy:

According to Einstein's mass-energy relationship ($E = mc^2$), in all nuclear reactions, the total sum of mass and energy must conserve. This means the total mass on the left side of the equation (before the decay) is equal to the total mass on the right side (after the decay), plus the Q value.

The Q value represents the mass difference before and after the decay, and this difference is converted into energy according to the mass-energy relationship ($E=mc^2$). This released energy is typically in the form of kinetic energy of the decay products, such as alpha particles, beta particles, or gamma rays.

So, the equation that represents the conservation of mass and energy in radioactive decay can be expressed as:

Total mass before decay=Total mass after decay+Q-value

This equation underscores the principle that while mass may appear to decrease due to the emission of particles; the lost mass is accounted for in the form of energy released during the decay process.

27.4.4 Nuclear Fission and Fusion:

The uranium nucleus, upon absorbing a neutron, underwent a process of splitting into two roughly equal parts. This revelation was striking because, up until then, known nuclear reactions had only involved the ejection of tiny fragments, such as neutrons, protons, or alpha particles, from a nucleus.

Nuclear Fission:

The splitting of a heavy nucleus (A > 230) into two medium-mass nuclei in a nuclear reaction with the release a huge amount of energy due to mass defect is called **nuclear fission**. For example, when a uranium nucleus (U-235) is bombarded by a slow moving neutron (called thermal neutron), the U-235 nucleus splits into two medium-mass nuclei with the release of huge amount of energy as shown in figure 27.7.

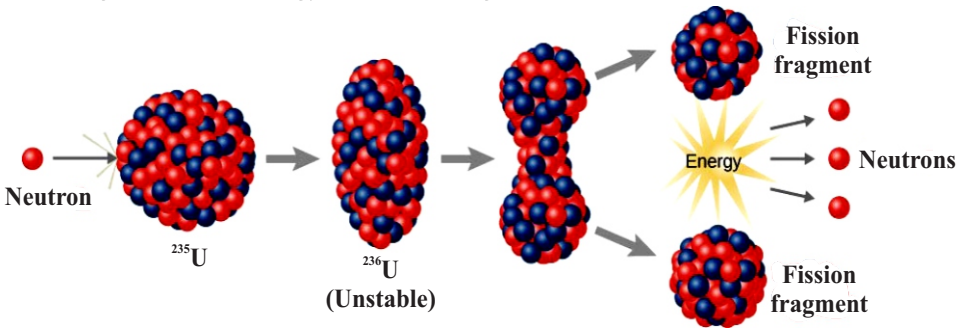


Figure 27.7 Heavy nuclei splits into two light nuclides (nuclear fission process)

This fission reaction is given below:

$$^{235}_{92}U + ^{1}_{0}n \rightarrow ^{236}_{92}U^{*} \rightarrow ^{141}_{56}Ba + ^{92}_{36}Kr + 3^{1}_{0}n + 200MeV \ldots\ldots\ldots 27.25$$

Where * on $\mathbf{U}^{*}$ represents that U-236 is in excited state. Two things are worth noting in this fission reaction. First, a huge amount of energy (about 200 MeV per U -235 nucleus) is released in the process.

Secondly, on the average 2 to 3 neutrons are released in the process. The released neutrons can further cause splitting of $_{92}U^{235}$ nuclei and lead to self-sustaining nuclear fission.

When nuclear fission takes place, it is found that the sum of the masses of fission products is very slightly less than the sum of the masses of reactant products. As a result, there occurs a mass defect (m) in nuclear fission. This mass defect is converted into energy according to the relation $E = mc^2$. The energy released in the above fission can be determined from the mass defect that occurs in the process.

Total mass before fission:

mass of U-235 = 235.043933 a.m.u
mass of neutron = 1.008665 a.m.u
sum of masses before fission = 236.052598 a.m.u

Total mass after fission:

mass of two fragments = 232.812000 a.m.u
mass of 3 neutrons = 3.025995 a.m.u
sum of masses after fission = 235.837995 a.m.u
Mass defect Δm= 236.052598-235.837995 = 0.214603 a.m.u
Therefore, energy released per fission of U-235 = 0.214603 X931.5 =**200 MeV**

Nuclear Fusion:

The process of nuclear fusion is just the reverse of nuclear fission. But the energy released per unit mass in nuclear fusion is much greater than the energy released per unit mass in nuclear fission. When two light nuclei are combined to form a heavy nucleus, the mass of the product nucleus is slightly less than the sum of the masses of the light nuclei fusing together.

The process of combining two light nuclei to form a heavy nucleus with the release of huge amount of energy due to mass defect is known as nuclear fusion.

This mass defect results in the release of a huge amount of energy according to the relation $E = mc^2$. When two nuclei of heavy hydrogen or deuterium ($_1H^2$) are combined, the following reaction is possible:

$$_1^2H + {}_1^2H \rightarrow {}_1^3H + {}_1^2H + 4MeV \ldots\ldots 27.26$$

figure 27.8 shows the nucleus of tritium ($_1H^3$) so formed can again fuse with a deuterium nucleus ($_1H^2$) to give the following reaction:

$$_1^3H + {}_1^2H \rightarrow {}_2^4H + {}_0^1n + 17.6MeV \ldots\ldots\ldots 27.27$$

The net result of these two nuclear reactions is that three deuterium ($_1H^2$) nuclei fuse together to form a helium nucleus ($_2He^4$) and a neutron with the release of 21.6 MeV (4.0 + 17.6 = 21.6 MeV). This energy of 21.6 MeV is obtained in

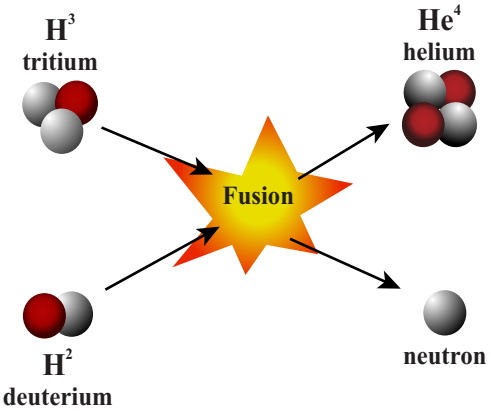


Figure 27.8 process of nuclear fusion

the form of kinetic energy of proton ($_1H^1$) and a neutron $_0n^1$.

Note that energy released in the fusion reaction is 21.6 MeV which is very much less than the energy of about 200 MeV released in the fission of $_{92}U^{235}$ nucleus. But this does not mean that fusion is a weaker energy source than fission. The sun and other stars are very hot so nuclei are moving fast enough for fusion to take place and the energy released keeps the temperature high so that further fusion reactions can occur. But on earth, such high temperatures are not attained in a controlled manner. However, the temperature produced by a fission bomb (atom bomb) is close to 10^8K. Therefore, **fission bomb** can be used to cause the fusion process.

The practical problems involved in producing energy from fusion to make a practical and cost-effective form of power is: Temperature, Pressure and confinement.

27.4.5 Fission Chain Reaction:

The nuclear fission which once started continues till all the atoms of the fissionable material are disintegrated is called chain reaction. Figure 27.9 shows how chain reaction is initiated in uranium by a single neutron. When a single neutron initially causes the fission of $_{92}U^{235}$ nucleus, 3 neutrons are released along with huge amount of energy. These 3 neutrons in turn cause three more nuclei to split, thereby liberating a total of 9 neutrons and so on. The process proceeds very quickly and in a very short time the whole of the uranium undergoes fission. For the fission to be self-sustaining, the number of emitted fission neutrons should be more than the incident ones. Under such conditions, the fission neutrons keep on increasing, thus maintaining the chain reaction. A very common term, called **neutron multiplication factor** (or **reproduction factor**), is often used in chain reaction. It is represented by k:

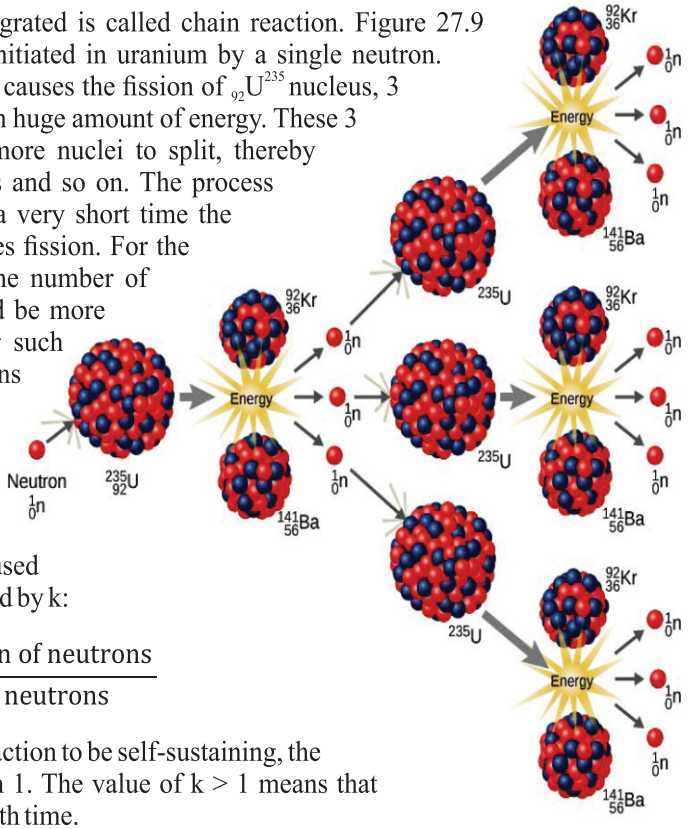


Figure 27.9 process of nuclear fission chain reaction

$$k = \frac{\text{rate of production of neutrons}}{\text{rate of loss of neutrons}}$$

Obviously, for chain reaction to be self-sustaining, the value of k must be greater than 1. The value of k > 1 means that neutrons increase or multiply with time.

There are two types of fission chain reactions (according to neutron multiplication factor k):

1. Controlled chain reaction:

A controlled chain reaction is a chain reaction in which the number of neutrons produced can be controlled. This allows for a sustained release of energy, which can be used for beneficial purposes, such as generating electricity. A chain reaction can be controlled by systematically removing some of the fission neutrons from the reaction vessel. The apparatus in which controlled chain reaction takes place is called a **nuclear reactor**.

DO YOU KNOW?

The atomic bombs dropped on Hiroshima and Nagasaki during World War II utilized uncontrolled chain reactions.

2. Uncontrolled chain reaction:

An uncontrolled chain reaction is a chain reaction in which the number of neutrons produced cannot be controlled. This results in a sudden and rapid release of energy, which can be destructive. For example, atomic bomb.

27.4.6 Nuclear Reactor:

A nuclear reactor is a device in which controlled fission chain reaction takes place. A nuclear reactor is also known as nuclear pile or atomic pile.

Such a system was first achieved with uranium as the fuel in 1942 by Enrico Fermi.He used uranium-235 isotope that releases energy through nuclear fission. The schematic diagram of nuclear reactor is shown in figure 27.10

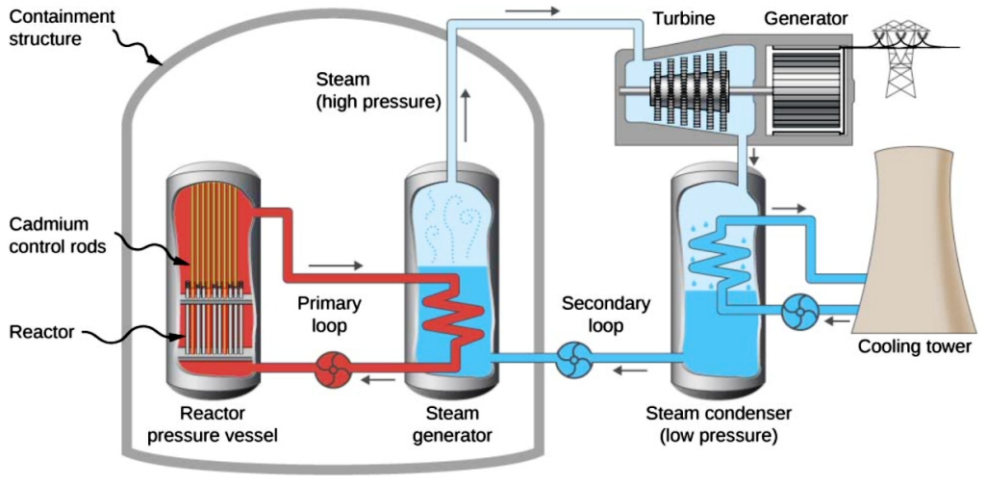


Figure 27.10 schematic diagram of Nuclear reactor

The following are the main components of the nuclear reactor:

Fissionable Substance:

Nuclear reactors use fuel, typically enriched uranium or plutonium, to sustain the fission chain reaction. The U-235 is fissionable, but uranium from ore typically contains only about 0.7 percent of U-235, with the remaining 99.3 percent being the U-238 isotope. Because

uranium-238 tends to absorb neutrons, reactor fuels must be processed to increase the proportion of U-235 so that the reaction can sustain itself. This process is called **enrichment**.

Moderator:

The function of the moderator is to slow down the highly energetic neutrons produced in the process of fission of U-235 to thermal energies. Heavy water (D_2O), graphite, beryllium, etc., are used as moderators. Ideally, moderators have low atomic weight and low absorption cross-section for neutrons.

Control Rods:

Control rods are made of materials like boron or cadmium that absorb neutrons, regulating the rate of the fission chain reaction. By adjusting the position of control rods within the reactor core, operators can control the power output and maintain stability.

Coolant:

Coolant circulates through the reactor core to transfer heat away from the fuel and other reactor components. Common coolants include water, heavy water, or gases like helium or carbon dioxide. The heated coolant then transfers its thermal energy to a secondary loop containing water, which turns into steam. This steam drives turbines connected to generators, producing electricity.

Protective Shield:

In a nuclear reactor, there are many types of harmful radiations emitted which are dangerous for all living things. In order to protect from these radiations, the reactor is surrounded by a massive biological shield.

27.4.7 Nuclear Fusion in Sun and Stars:

Every second, the sun fuses around 500 million metric tons of hydrogen in its core. The core of the sun is incredibly hot, with temperatures reaching about 20 million degrees Celsius, while its surface temperature is around 5 million degrees Celsius.

The sun is a star which is primarily made up of hydrogen (about 75%), helium (about 25%), and trace amounts of other elements. It produces energy through a process called nuclear fusion, where hydrogen atoms combine to form helium atoms, releasing light and heat in the process.

The fusion in the sun can take place in two different reaction sequences, the most common of which, the **Proton-Proton (PP) Cycle** and the other one is **Carbon-Nitrogen-Oxygen (CNO) Cycle**.

The Proton-Proton Cycle involves the fusion of protons (hydrogen nuclei) to form helium as shown in figure 27.11. Hans Bethe was the first to work out the detailed steps of the PP cycle in 1938. The PP cycle is a

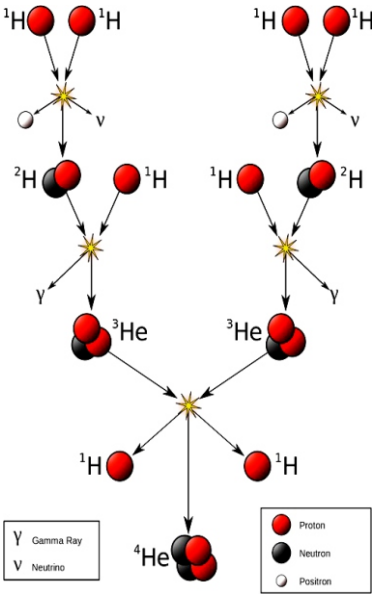


Figure 27.11
Proton-proton cycle to form helium

very efficient way to generate energy in the Sun. In the p–p chain, two protons first fuse to produce a deuterium nucleus which combines with another proton to yield ${}^{3}_{2}He$. Two ${}^{3}_{2}He$ nuclei fuse and form ${}^{4}_{2}He$ and two protons. These reactions can be represented by the equations

$$ {}^{1}_{1}H + {}^{1}_{1}H \rightarrow {}^{2}_{1}H + {}^{0}_{1}e + \nu + Q $$

$$ {}^{1}_{1}H + {}^{2}_{1}H \rightarrow {}^{3}_{2}He + \gamma + Q $$

$$ {}^{3}_{2}He + {}^{3}_{2}He \rightarrow {}^{4}_{2}He + {}^{1}_{1}H + {}^{1}_{1}H + Q \ldots\ldots\ldots 27.28 $$

The net Q value of the chain reactionsisabout 26 MeV.

The Carbon-Nitrogen-Oxygen (CNO) cycle was independently suggested by Carl von Weizsacker and Hans Bethe in the late 1930s. The CNO cycle is a series of nuclear fusion reactions that convert hydrogen into helium as predicted in figure 27.12, and it is the primary source of energy in stars that are more than 1.3 times as massive as the Sun. This cycle uses carbon, nitrogen, and oxygen as catalysts to convert hydrogen to helium:

$$ {}^{12}_{6}C + {}^{1}_{1}H \rightarrow {}^{13}_{7}N + \gamma $$

$$ {}^{13}_{7}N \rightarrow {}^{13}_{6}C + e^{+} + \nu_e $$

$$ {}^{13}_{6}C + {}^{1}_{1}H \rightarrow {}^{14}_{7}N + \gamma $$

$$ {}^{14}_{7}N + {}^{1}_{1}H \rightarrow {}^{15}_{8}O + \gamma $$

$$ {}^{15}_{8}O \rightarrow {}^{15}_{7}N + e^{+} + \nu_e $$

$$ {}^{15}_{7}N + {}^{1}_{1}H \rightarrow {}^{12}_{6}C + {}^{4}_{2}He + \gamma $$

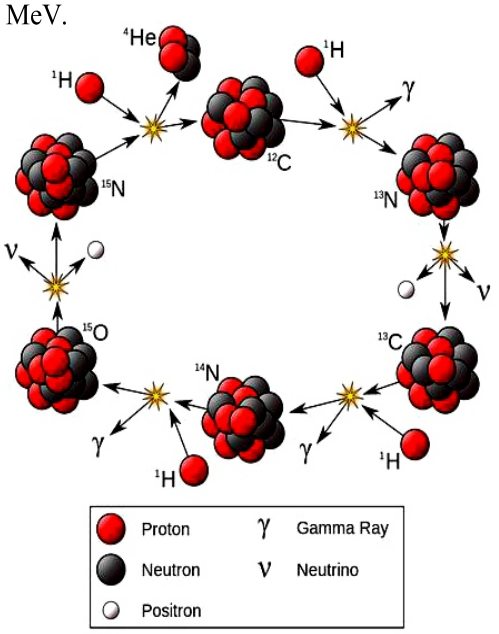


Figure 27.12
The Carbon-Nitrogen-Oxygen (CNO) cycle

In the CNO cycle, four protons fuse, using carbon, nitrogen and oxygen isotopes as a catalyst, to produce one alpha particle, two positrons and two electron neutrinos. Combining all the above reactions, the net reaction for the CNO cycle comes out to be

$$ 4\,{}^{1}_{1}H \rightarrow {}^{4}_{2}He + 2\beta^{+} + 2\vartheta + 26.7MeV \ldots\ldots\ldots 27.29 $$

We see that the net energy release is nearly same for both cycles.

27.5 Radiation Exposure:

Radiation refers to the transmission of energy in the form of waves or particles. This energy spectrum contains a vast range, from gentle radio waves and visible light to the more impactful X-rays, gamma rays, and even cosmic rays. The key factor influencing its impact is the **dose**, measured in units like milli-sieverts (mSv).

27.5.1 Nuclear Radiation Exposure and its Biological Effects:

Exposure is defined as the amount of ionization produced in a unit mass of dry air at standard pressure (STP). Its unit is 1 roentgen = 1R = 2.58 X 10^{-4} C/kg.

There are two main types of radiation exposure: **external** and **internal**.

1. **External radiation** exposure occurs when a person is exposed to radiation from a source outside the body, such as an X-ray machine or a nuclear power plant.
2. **Internal radiation** exposure occurs when a person ingests or inhales radioactive material, which can then accumulate in the body and emit radiation.

Radiation exposure can also be measured in other units, including sieverts (Sv), millisieverts (mSv), and microsieverts (μSv).

Biological Effects of Radiation:

When radiation is absorbed by matter, especially living tissue, it can induce significant changes. The biological effects of radiation are diverse and depend on several factors, including:

- **Type of radiation:** Each type interacts differently. Alpha particles, while unable to penetrate deeply, cause disorder within cells they reach. Beta particles travel farther but deposit less energy, while gamma rays can pierce through tissues, potentially affecting multiple organs.
- **Dose absorbed:** This quantifies the energy deposited per unit mass of tissue. Higher doses generally translate to more pronounced effects.
- **Duration of exposure:** Acute (short-term) exposure like an X-ray differs from chronic (long-term) exposure from environmental sources or occupational hazards.
- **Individual sensitivity:** Age, health status, and genetic predispositions can influence susceptibility.

The biological effects of radiation include:

- **Acute radiation syndrome (ARS):** High-dose exposure can trigger ARS, a complex condition affecting multiple organ systems with symptoms like nausea, vomiting, hair loss, and bone marrow suppression.
- **Cancer risk:** Chronic low-dose or high-dose exposures can increase the risk of various cancers, depending on the affected tissue.
- **Genetic effects:** Germ cell mutations can lead to congenital disabilities in offspring.
- **Reproductive issues:** Fertility can be impaired, and miscarriage risk may increase.
- **Developmental effects:** Early exposure inutero can affect fetal development, causing physical and cognitive impairments.

27.5.2 Dosimetry:

Dosimetry is the science of measuring and quantifying radiation dose. It is a field of physics that deals with the interaction of radiation with matter and the biological effects of radiation exposure.

Dosimetry is derived from the Greek word ‘dos’, meaning ‘gift’ or ‘a giving’. In the context of physics, it represents the scientific discipline that quantifies the amount of radiation ‘given’ to or absorbed by an object or body. The goal is to determine potential biological effects, ensuring safe limits and monitoring radiation exposure.

Key Components of Dosimetry:

1. **Absorbed Dose:** The amount of radiation energy absorbed per unit mass of the material. Measured in Gray (Gy), it is a pivotal metric in understanding radiation’s effect on tissues.

2. **Equivalent Dose:** Not all radiation types have the same biological impact, even if their absorbed doses are identical. To account for this, radiation weighting factors are introduced to derive the equivalent dose. Measured in Sievert (Sv), it provides a more biologically relevant dose metric.
3. **Effective Dose:** Considering that different tissues have varying sensitivities to radiation, tissue weighting factors are used. The effective dose, also in Sieverts (Sv), offers a measure that summarizes the potential overall harm to the whole organism.
4. **Operational Dose Quantities:** These are practical quantities used for routine monitoring in radiological protection. They are designed to estimate effective dose or equivalent dose to a particular tissue. Examples include ambient dose equivalent and personal dose equivalent.

Dosimetry is performed using devices known as **dosimeters**, which can be worn by people working with or around radioactive materials to monitor their exposure levels. These devices can measure and record the dose of radiation over time.

27.5.3 Medical Uses of Nuclear Radiation:

Medical use of nuclear radiation is quite common in today's hospitals and clinics. It contains various diagnostic and therapeutic applications that control the properties of ionizing radiation to diagnose and treat diseases. Some examples include:

Imaging:

- **X-rays:** The workhorse of medical imaging, X-rays utilize electromagnetic radiation to reveal fractures, bone structures, and internal injuries.
- **CT scans:** Combining multiple X-ray images, CT scans provide detailed cross-sectional views of organs and tissues, aiding in diagnosing tumors, infect Ions, and internal bleeding.
- **PET scans:** Positron emission tomography (PET) utilizes radioactive tracers to map metabolic activity within the body, uncovering cancerous tumors and other metabolic disorders.

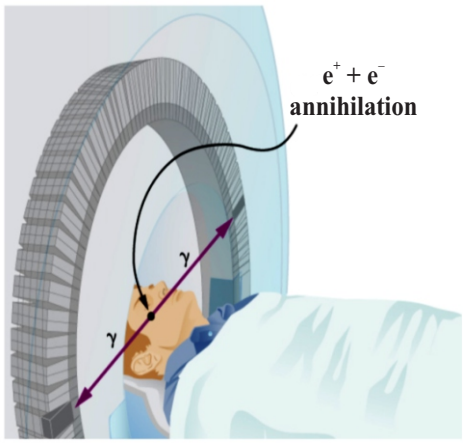


Figure 27.13 The CT scan

Nuclear medicine procedures:

- **Bone scans:**
  Radioactive tracers identify bone diseases like osteoporosis and cancer metastases.
- **Thyroid scans:**
  Radioactive iodine helps diagnose and treat thyroid disorders.
- **Lung scans:**
  Technetium-99m helps assess lung function and detect blood clots.

Treatment:

- **Radiotherapy:** Harnessing the destructive power of radiation, targeted beams are used to shrink and destroy cancerous tumors with remarkable precision. This non-invasive therapy plays a crucial role in treating various cancers, often in conjunction with surgery and chemotherapy.
- **Brachy therapy:** Radioactive implants placed directly within tumors deliver high doses of radiation locally, minimizing damage to surrounding tissue. This approach is particularly effective for treating prostate, cervical, and head and neck cancers.
- **Radioisotope therapy:** Radioactive isotopes are used to treat specific conditions like hyperthyroidism (radioactive iodine) and blood disorders (radioactive phosphorus).

Other Applications:

- **Pain management:** Radiofrequency ablation utilizes radio waves to heat and destroy nerve tissue, offering pain relief for chronic conditions like back pain and arthritis.
- **Sterilization:** Medical instruments and equipment are sterilized using gamma radiation, ensuring sterility and preventing infections.

27.5.4 Exposure of radiation:

This process creates charged particles (Ions) and free radicals, which can disrupt molecular structures and biological processes. Examples of ionizing radiation include:

- **X-rays:** Electromagnetic radiation produced by high-energy electron beams or X-ray tubes. X-rays are commonly used in medical imaging, security screening, and industrial applications.
- **Gamma rays:** High-energy electromagnetic radiation emitted by radioactive decay processes, such as those occurring in radioactive isotopes like cobalt-60 and cesium-137. Gamma rays are used in medical imaging (gamma cameras), radiation therapy, and sterilization processes.
- **Alpha particles:** Alpha particles are emitted during the decay of heavy elements like uranium and radium and have low penetrating power but can cause significant damage if inhaled or ingested.
- **Beta particles:** Beta particles can penetrate deeper into tissues than alpha particles but are less damaging. They are used in medical imaging (positron emission tomography) and radiation therapy.
- **Neutrons:** Neutrons can induce nuclear reactions and are used in neutron activation analysis, neutron radiography, and certain types of cancer therapy.
- **Cosmic rays:** High-energy particles from outer space continuously bombard Earth's atmosphere, generating secondary radiation that reaches the surface.
- **Industrial applications:** Gauges, sterilization processes, and smoke detectors often employ ionizing radiation for various industrial purposes.

These forms of ionizing radiation have various applications in medicine, industry, research, and other fields but require careful handling and monitoring to minimize risks to human health and the environment.

27.5.5 Uses of Radioactive Tracers:

One of the most important uses of nuclear radiation is the location and study of diseased tissue. This can be done by radioactive tracers. **Radioactive tracers** are radioactive substances that are used to track the movement and behavior of specific molecules or compounds in various systems. These tracers emit radiation that can be detected through biochemical reactions, metabolic pathways, fluid flow, and environmental transport. Some examples of the uses of radioactive tracers include:

Medical Imaging:

- **Positron Emission Tomography (PET):** Radiotracers labeled with positron-emitting isotopes (e.g., fluorine-18, carbon-11) are injected into the body and used to visualize metabolic activity, blood flow, and receptor binding in tissues. PET imaging is valuable for diagnosing and monitoring diseases such as cancer, heart disease, and neurological disorders.

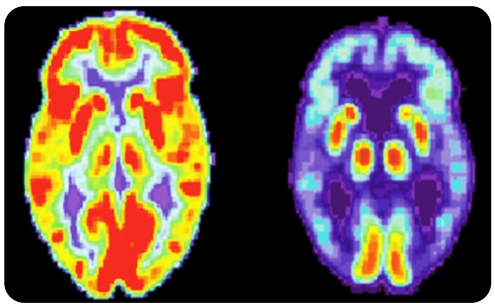


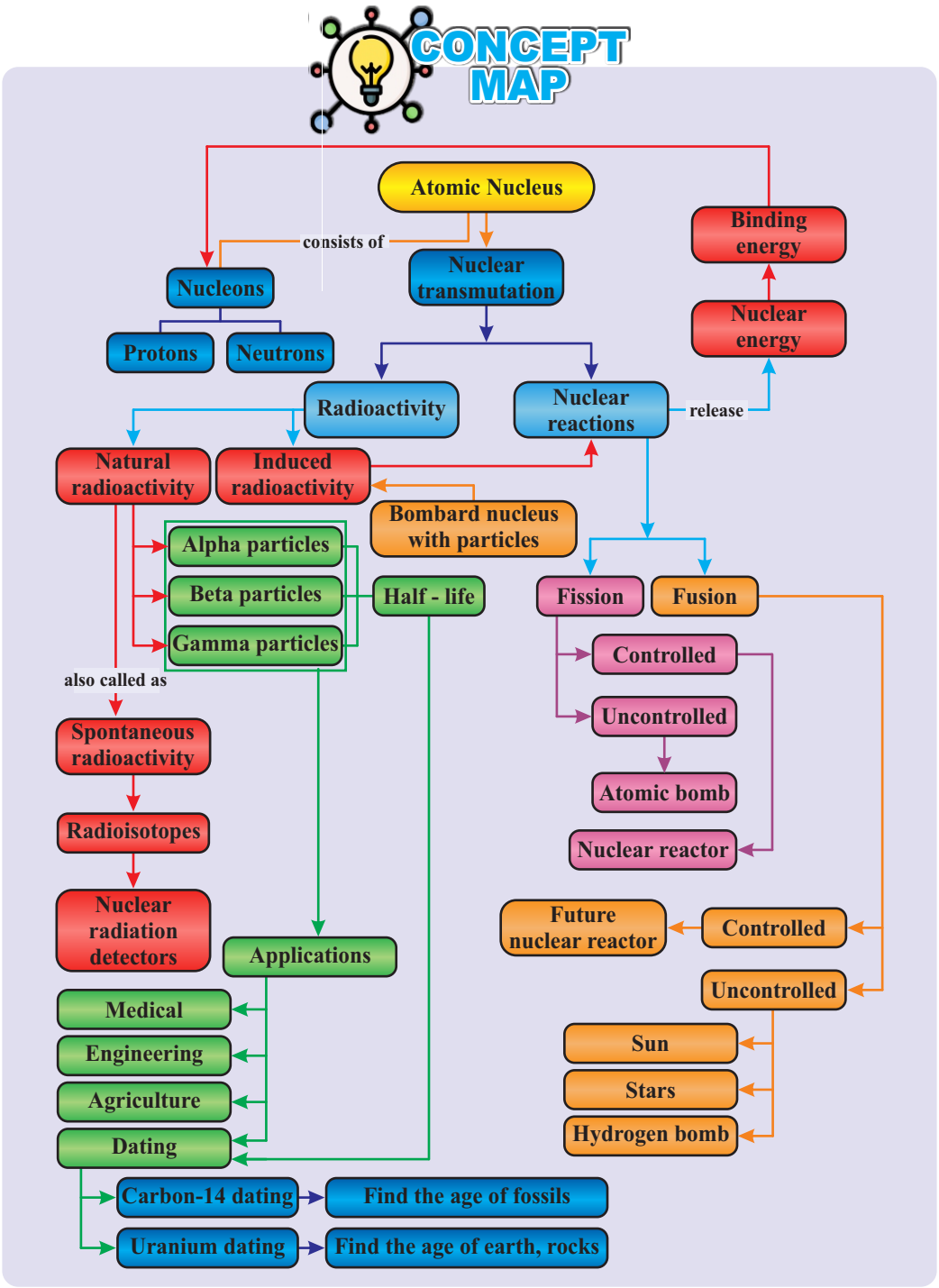
Figure 27.14 Positron Emission Tomography

Industrial Processes:

- **Flow Visualization:** Radiotracers are injected into fluids or gases to track flow patterns and detect leaks or blockages in industrial pipelines, heat exchangers, and reactors. This technique is used in industries such as petrochemicals, food processing, and nuclear power.
- **Process Optimization:** Radiotracers are used to monitor and optimize chemical reactions, mixing processes, and material transport in industrial processes. By tracking the movement of tracers, engineers can identify bottlenecks, improve efficiency, and ensure product quality.
- **Soil Fertility and Nutrient Uptake:** Radioactive isotopes like phosphorus-32 (^{32}P) are used to study nutrient uptake in plants. By tagging fertilizers with these tracers, scientists can track how nutrients move through the soil and are absorbed by plants. This information helps in optimizing fertilizer application, ensuring that crops receive the right amount of nutrients.
- **Environmental Impact Assessments:** Radioactive tracers help in studying the environmental impact of agricultural practices. For instance, they can be used to trace the movement of pollutants or contaminants from agricultural fields to water bodies, enabling better management practices to protect the environment.



- ✓ Variants of a particular chemical element that have the same number of protons but different numbers of neutrons, resulting in different atomic masses is known as isotopes.
- ✓ Mass Spectrograph (Analytical device) used to measure the charge to mass ratio of ions, useful in identifying isotopic compositions and molecular structures.
- ✓ The process (radioactivity decay) by which an unstable atomic nucleus loses energy by emitting radiation, resulting in the transformation of one element into another
- ✓ Activity measures the rate of decay of a radioactive substance, while the decay constant is a probability factor that describes the likelihood of an atom decaying per unit time.
- ✓ The time required for half of the radioactive atoms in a sample to decay, a characteristic property of each radioactive isotope.
- ✓ The energy (Binding Energy) required disassembling a nucleus into its individual protons and neutrons, indicating the stability of the nucleus.
- ✓ Fission is the splitting of a heavy nucleus into lighter nuclei with the release of energy, while fusion is the combining of light nuclei to form a heavier nucleus, also releasing energy.
- ✓ A self-sustaining sequence of nuclear fission reactions, where the neutrons produced by each fission event causes further fissions is known as chain reaction
- ✓ A device (Nuclear Reactor) used to initiate and control a sustained nuclear chain reaction, commonly used for energy production.
- ✓ The absorption (Nuclear Radiation Exposure) of energy from nuclear radiation by living tissues, which can cause cellular damage and increase cancer risk.
- ✓ Dosimetry is the measurement and calculation of the absorbed dose of radiation by the human body, crucial for assessing exposure and ensuring safety.
- ✓ Strategies include time minimization, distance maximization, and shielding, along with regulatory measures to protect individuals from harmful radiation levels is called Limiting exposure to ionizing radiation.


CONCEPT MAP
Atomic Nucleus
consists of
Nucleons
Protons
Neutrons
Nuclear transmutation
Radioactivity
Nuclear reactions
release
Binding energy
Nuclear energy
Natural radioactivity
Induced radioactivity
Bombard nucleus with particles
Alpha particles
Beta particles
Gamma particles
Half - life
also called as
Spontaneous radioactivity
Radioisotopes
Nuclear radiation detectors
Applications
Medical
Engineering
Agriculture
Dating
Carbon-14 dating
Find the age of fossils
Uranium dating
Find the age of earth, rocks
Fission
Fusion
Controlled
Uncontrolled
Atomic bomb
Nuclear reactor
Future nuclear reactor
Controlled
Uncontrolled
Sun
Stars
Hydrogen bomb

EXERCISE

Section (A): Multiple Choice Questions (MCQs)

Choose the correct answer:

1. Identify particle x in the following nuclear reaction ${}^{9}_{4}Be + {}^{4}_{2}He \rightarrow {}^{12}_{6}C + x$
   (a) Electron (b) proton (c) neutron (d) photon
2. In the equation ${}^{27}_{13}Al + {}^{4}_{2}He \rightarrow {}^{30}_{15}P + X$,The correct symbol for X is:
   (a) ${}^{-1}_{0}e$ (b) ${}^{1}_{1}H$ (c) ${}^{4}_{2}He$ (d) ${}^{1}_{0}n$
3. Beta rays emitted by a radioactive material are:
   (a) Electromagnetic radiations (b) The electrons orbiting around the nucleus
   (c) Charged particles emitted by the nucleus (d) Neutral particles
4. The number of α and β particles emitted in the following radioactive decay is:
   ${}^{200}_{90}X \rightarrow {}^{168}_{80}Y + {}^{?}_{?}X$
   (a) 8 and 6 (b) 6 and 8 (c) 8 and 8 (d) 6 and 6
5. If radium has half-life of 5 years. Thus for a nucleus in a sample of radium, the probability of decay in ten years is:
   (a) 50% (b) 75% (c) 100% (d) 60%
6. Emission of β particles by an element affects its mass number **A** in the following way:
   (a) Increases by 1 (b) decreases by 1
   (c) Increases by 2 (d) remains the same
7. As the number of nucleons in a nucleus increases, the binding energy per nucleon:
   (a) Increases continuously with mass number
   (b) Decreases continuously with mass number
   (c) remains constant with mass number
   (d) First increases and then decreases with increase of mass number
8. Moderator in a nuclear reactor slows down the neutrons to:
   (a) Decrease the probability of escape (b) Increase the probability of nuclear fission
   (c) Decrease the probability of absorption (d) all of the above
9. Emission of β^{+}particles by an element affects its atomic number Z in the following way:
   (a) Increases by 1 (b) decreases by 1
   (c) Increases by 2 (d) remains the same
10. The half life period of a radioactive element is 100 days. After 400 days, 16 g of the element will be reduced to
    (a) 8g (b) 4g (c) 2g (d) 1g

Section (B): CRQs (Short Answered Questions):

1. Why are protons and neutrons necessary for the stability of an atomic nucleus?
2. How do isotopes of an element differ and why are these differences significant?
3. Name the two methods of controlling a chain reaction.
4. Why is nuclear decay described as spontaneous and random?
5. Explain what happens during the nuclear fusion process.

6. How are activity and decay constant related in radioactive materials?
7. What is meant by binding energy and binding energy per nucleon?
8. Why is the concept of half-life important in studying radioactive substances?
9. Explain the process of positron emission.

Section (C): ERQs (Long Answered Questions):

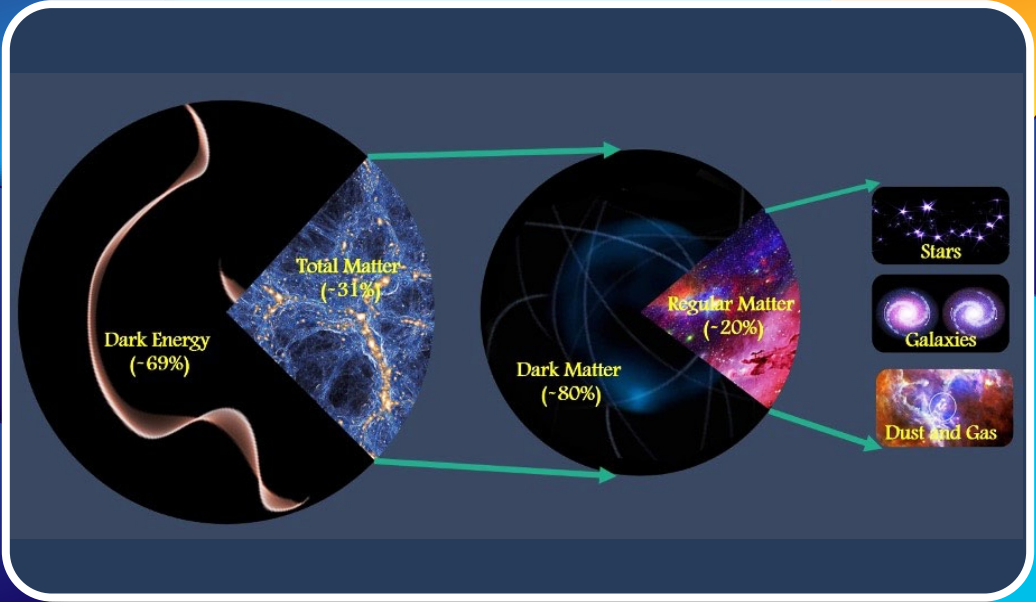
1. What is radioactivity? State the law of radioactive disintegration. Show that radioactive decay is exponential in nature.
2. Explain the main differences between alpha, beta, and gamma emissions.
3. A fission reactor produces energy to drive a generator. Describe briefly how this energy is produced.
4. Define Q-value of a nuclear reaction and its significance.
5. Describe the construction and working of GM counter.
6. Discus nuclear reactions induced by neutrons. Why are neutrons preferred to other particles?
7. What is a nuclear reaction? Explain nuclear fission and fusion.
8. How do the Sun and stars produce energy? What is the proton-proton cycle? Explain with details.

Section (D): Numerical:

1. In 9.0 days the number of radioactive nuclei decreases to one-eighth the number present initially. What is the half-life (in days) of the material? **[Ans: 3 days]**
2. The $^{32}_{15}P$ isotope of phosphorus has a half-life of 14.28 days. What is its decay constant in units of s^{-1}? **[Ans: $5.62\times10^{-7}s^{-1}$]**
3. Find the binding energy (in MeV) for lithium $^{7}_{3}Li$ (atomic mass = 7.016 003 u). **[Ans: 39.2 MeV]**
4. The binding energy of a nucleus is 225.0 MeV. What is the mass defect of the nucleus in atomic mass units? **[Ans: 0.2415 u]**
5. A copper penny has a mass of 3.0 g. Determine the energy (in MeV) that would be required to break all the copper nuclei into their constituent protons and neutrons. Ignore the energy that binds the electrons to the nucleus and the energy that binds one atom to another in the structure of the metal. For simplicity, assume that all the copper nuclei are $^{63}_{29}Cu$ (atomic mass = 62.939 598 u). **[Ans: 1.51249×10^{25}MeV]**
6. Write the β^+ decay process for each of the following nuclei with their proper chemical symbols including Z and A for each daughter nucleus: (a) $^{18}_{9}F$(b) $^{15}_{8}O$.
   [Ans: (a) $^{18}_{9}F \rightarrow {}^{18}_{8}O + e^+ + \nu_e$ (b) $^{15}_{8}O \rightarrow {}^{15}_{7}N + e^+ + \nu_e$]
7. A device used in radiation therapy for cancer contains 0.50 g of cobalt $^{60}_{27}Co$ (59.933 819 u). The half-life of $^{60}_{27}Co$ is 5.27 years. Determine the activity of the radioactive material.
   [Ans: 2.10×10^{13}decays/second (Bq)]

Unit 28 Particle Physics

Teaching Periods 03 **Weightage %** 02



In particle physics, Dark Matter and Dark Energy are mysterious entities that make up about 95% of the universe's mass-energy budget. Dark Matter is thought to be composed of Weakly Interacting Massive Particles (WIMPs), such as axions or neutralinos, which interact with normal matter only through gravity and the weak nuclear force, making them invisible to most detectors. Dark Energy, on the other hand, is believed to be a property of space itself, represented by a hypothetical field, like the Higgs field, that drives the accelerating expansion of the universe. Together, they form the "dark sector" of the universe, governing its large-scale structure and evolution, yet remaining elusive and invisible to direct detection, sparking intense research and debate in the particle physics community.

In this unit student should be able to:

- Describe the fundamental forces of nature and their field particles
- Describe the key features and components of the standard model of matter including hadrons, leptons and quarks.
- Describe the working principal, construction and use of Wilson Cloud Chambers
- Describe the working principal, construction and use of GM counter

Introduction:

Let's explore how our universe was formed and what holds everything together. People have been curious about these questions for a very long time. Originally, scientists believed that atoms were the smallest pieces that made up everything. But, in the 1800s, they discovered that atoms have smaller parts called electrons, protons, and neutrons.

Today, physics tells us about even tinier pieces that make up everything. These super small pieces are called "elementary" or "fundamental" particles, and they're so simple that they don't have any smaller parts inside them. Molecules, atoms, protons, and neutrons are not fundamental particles because these are made up of other particles, while electrons, quarks, and photons are fundamental particles. Studying all these elementary particles and understanding how they behave is the main goal of particle physics. Imagine these particles as tiny puzzle pieces that collectively form everything around us! To make sense of this cosmic puzzle, scientists have developed various theories. These theories aid in explaining the behavior and interactions of these particles. In this chapter, we will study one of the most crucial theories known as the **Standard Model**.

DO YOU KNOW?

Fundamental particles
Those particles which are simple, structure less, and are not made up of other particles.

Particle Physics is the branch of Modern Physics which deals with the study and search of the ultimate constituents of matter (elementary particles) and their interactions.

28.1 The Standard Model:

Our universe is made up of two things: matter and energy (radiation). To understand them better, scientists have divided particles into two main groups: matter particles and force particles. Scientists have identified many elementary particles belonging to these categories. These particles are categorized and explained in detail in the Standard Model of Particle Physics, which is the best-known theory to date. It is a framework that explains three of the four fundamental forces (electromagnetism, the weak nuclear force, and the strong nuclear force) and all known elementary particles.

The Standard Model classifies all known elementary particles into two main classes:

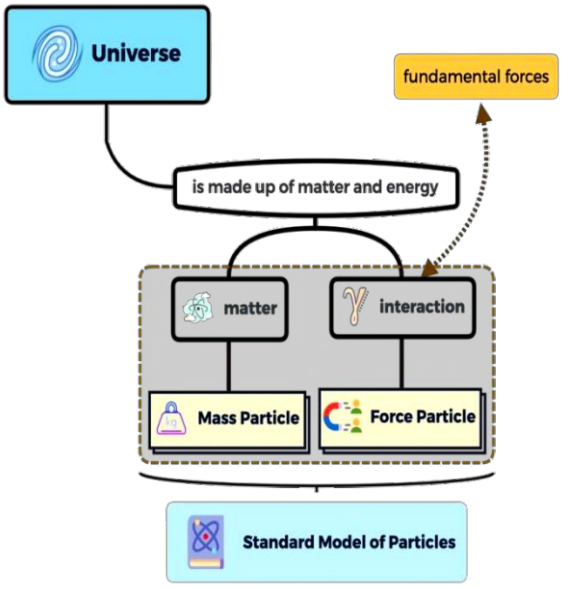


Figure 28.1:
The universe is made up of matter and energy.

Fermions:

These are the matter particles, which make up everything in the universe.

Bosons:

These are the force-carrying particles, which mediate the interactions between fermions.

The Standard Model is a very successful theory. It has been able to predict the existence of many new particles, and it has been confirmed by a wide range of experiments. However, the Standard Model is not perfect. It does not explain gravity, which is the fourth fundamental force. It also does not explain dark matter and dark energy, which make up most of the universe.

28.1.1 Fundamental forces and their field particles:

The building block or brick of all matter is known as an elementary particle, but these particles cannot make up the whole structure of the universe by themselves. Just like a "brick" is the building block of our homes, we need cement as "glue" to give structure to our homes. Similarly, we need "glue" or "force particles" to build our universe. This glue is known as fundamental forces in the Standard Model. Each fundamental force has a field particle associated with it. A field particle is a virtual particle that mediates the fundamental force between two other particles. The Standard Model has kept all field particles in a class of "Boson."

The term "field particles" refers to particles associated with force fields. According to the Standard Model, bosons are often considered field particles because they are linked to force fields. For example, the photon is a field particle associated with the electromagnetic field.

Scientists have grouped all fundamental forces into four basic types. In order of increasing strength, these forces and their associated field particles are described as:

1. Gravitational Force:

It is the force that attracts two masses towards each other. It is the weakest of the four fundamental forces but acts over long distances.

Field Particle: The hypothetical particle associated with gravity is the graviton, although it has not been observed yet. The Standard Model doesn't explain gravity.

2. Weak Nuclear Force:

The weak force is responsible for radioactive decay, where unstable atomic nuclei break down into smaller, more stable nuclei. It's responsible for processes like beta decay and neutrino emission. This force is weaker than electromagnetic and strong nuclear forces but stronger than gravitational force.

Field Particle: There are three field particles associates with weak nuclear force. These are W^+, W^- and Z bosons. These short-lived bosons carry the force over very small distances, explaining the limited range of the weak force.

3. Electromagnetic Force:

This force is responsible for the interactions between charged particles, such as electrons and protons. It includes both electric and magnetic forces. The electromagnetic force

is stronger than both gravitational and weak nuclear forces. It is also a long-range force, similar to the gravitational force.

Field Particle: The field particle of electromagnetic force is photon. It is mass less and chargeless particle. When ever charged particles interact, they exchange photons, causing the attractive or repulsive forces we observe.

4. Strong Nuclear Force:

The strong force binds quarks together to form protons and neutrons, and it holds atomic nuclei together. It is the strongest force among all forces and acts at subatomic levels.

Field Particle: Gluons are the field particles that mediate the strong force between quarks.

28.1.2 Particle zoo:

One way of studying elementary particles is to classify them into different categories based on certain behaviors and then to look for similarities or common characteristics among the classifications. We know that the Standard Model has classified all elementary particles and their interactions into two main groups according to their spins:

- Particles that carry force, called **boson s**having spin in integer values such as 0,1and 2.
- Particles that make up all matter, called **fermions** having spin in odd half integer values such as ½. 3/2 etc.

Bosons:

The arrangement of all elementary particles in the Standard Model constitutes a particle zoo. All bosons can be classified into two types: Elementary Boson and Composite Boson. Composite bosons consist of quark and anti-quark combinations while elementary bosons are not made up of other elementary particles.

Elementary Bosons:

Elementary bosons are the carriers of the fundamental forces in nature. For examples:

- Gluon
- W^+, W^-andZ^0 Bosons
- Photon
- Higgs Boson– *gives mass to all particles.*
- Graviton

The properties of elementary bosons are given in the table.

Table 28.1 The properties of elementary bosons

Force	Boson	Spin	Strength	Mass
Strong	Gluon	1	1	Massless
Electromagnetic	Photon	1	10^{-2}	Massless
Weak	W, Z	1	10^{-7}	80.91GeV
Gravity	Gravitation	2	10^{-39}	Massless

Composite Bosons:

Composite bosons are made up of an even number of fermions. For examples, **mesons** are composite bosons that composed of one quark and one antiquark. Mesons are intermediate mass particles that mediate the strong force between nucleons such as protons and neutrons.

Fermions:

All material particles are made up of fermions. We can further divide fermions in to two sub-classes: **elementary fermions** and **composite fermions**.

Elementary Fermions:

Elementary fermions are building blocks of all material particles. There are not made of any other particles. The fermions come in two types: **leptons** and **quarks**.

(a) Leptons:

Leptons are a group of elementary particles that do not experience the strong nuclear force. There are six types or flavors of leptons, which come in three pairs. The pairs are made up of three charged particles named electron, muon, and tau, along with their Partners called neutrinos (charge less). These all six leptons group into three generations:
Generation I, Generation II and Generation III as shown in the table. 28.2 Each generation consists of one pair of leptons (for example, electron with electron neutrino).
Properties of leptons:

- **Electron:** Negatively charged; commonly found in atoms.
- **Muon and Tau:** Heavier counter parts of the electron.
- **Neutrinos:** Electrically neutral; they interact very weakly with matter.
- Leptons interact via weak and electromagnetic forces but not through the strong force.
- Leptons are stable particles and do not undergo decay under normal circumstances.
- Leptons exist alone and donot form groups.

DO YOU KNOW?

- Our universe belongs to Generation I particles. Particles belonging to Generation II and Generation III are heavier than the particles belonging to Generation I.
- The heavier particles of Generations II and III undergo decay processes, transforming into Generation I particles.

Table 28.2 Generations of Leptons and Quarks

	Fermions			
	Charge	**Generation I**	**Generation II**	**Generation III**
Leptons	-1	electron (e)	muon (μ)	tau (τ)
	0	e-neutrino	μ-neutrino	τ-neutrino
Quarks	+2/3	up (u)	charm(c)	top(t)
	-1/3	down (d)	strange (s)	bottom(b)

Quarks:

Quarks are elementary particles that experience all three fundamental forces: strong nuclear force, weak nuclear force, and electromagnetic force. Quarks come in six types, or flavors: up (u), down (d), charm (c), strange (s), top (t), and bottom (b).

Quarks are a fundamental component of visible matter. All the matter around us, such as protons and neutrons, is made up of quarks. Similar to leptons, quarks also come in three generations, with each generation containing one pair of quarks. The up and down quarks are the only stable quarks in ordinary matter.

DO YOU KNOW?

Color force increases with increasing distance. The up and down quarks are the only stable quarks in ordinary matter.

Properties of Quarks:

- They carry fractional electric charges, either +2/3 or -1/3.
- Quarks interact strongly with the strong nuclear force, which is mediated by gluons. Quarks are never found as free particles in nature; they are always confined within larger particles called hadrons.
- Hadrons are particles made up of quarks held together by the strong nuclear force. Examples of hadrons include protons and neutrons.
- Quarks can undergo weak interactions, leading to processes such as beta decay. Weak interactions can change one type of quark into another. For example, a down quark can change into an up quark through weak decay processes.

DO YOU KNOW?

Hadron is Greek words means heavy. Meson has spin1, so they can also be placed in the class of boson.

Self-Assessment Questions:

1. Which particles don't participate in strong nuclear force?
2. Define the term "flavor" in the context of quarks and list the six quark flavors.

Color Charge:

1. Electric charge comes in only one type: positive (with its opposite negative). But strong charge (that deals with strong nuclear force) comes in three types: red, green, and blue. These color names are just labels and do not correspond to actual colors in the visual spectrum.
2. Quarks carry either one of the three color charges; and they can change their colors during particle interactions. Quarks of different colors are attracted to one another due to the strong nuclear force; it means red attracts green; blue attracts red, and so on. On the other hand, quarks of the same color repel one another.
3. Quarks always combine in ways that result in "color-neutral" or "white-color" particles. For example, a proton consists of three quarks: one red, one green, and one blue, making it color-neutral. Only white color combinations are permitted. This is why isolated quarks do not exist in nature. This is known as quark confinement. Therefore, all free particles have a color charge of zero.

Composite Fermions:

Composite subatomic particles such as protons, neutrons, alpha particles etc., are composed of two or more elementary particles. All composite particles are massive. The composite particles that are made of quarks are called hadrons. The two main categories of hadrons are baryons and mesons.

(a) Baryons:

Baryons are a class of hadrons that consist of three quarks. For example, protons and neutrons are the most well-known baryons, each composed of three quarks:

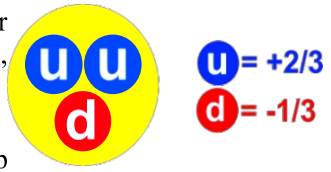


Figure 28.2 Proton Quarks

Proton:

It consists of two up and one down quarks ($\bar{u}\bar{u}\bar{d}$). Up (u) quark has +2/3 charge and down quark has -1/3charge. Therefore, net charge on proton has +1 charge:

$$Proton = \bar{u}\bar{u}\bar{d} = \frac{2}{3} + \frac{2}{3} - \frac{1}{3} = +1$$

$$Neutron = \bar{u}\bar{d}\bar{d} = \frac{2}{3} - \frac{1}{3} - \frac{1}{3} =$$

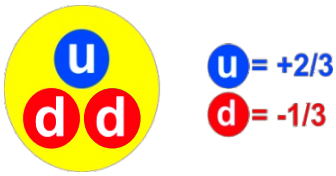


Figure 28.3 Neutron Quarks

Neutron:

It consists of three quarks: one up and two down quarks ($\bar{u}\bar{d}\bar{d}$). The net charge on neutron is zero:

(b) Mesons:

Mesons are another class of hadrons, but they consist of one quark and one anti quark. For example, pions (π) are common mesons. They carry the strong nuclear force, binding protons and neutrons together.

Quarks and Beta Decay:

β -decay is a type of radioactive decay process in which a neutron in an atomic nucleus is transformed into a proton. When a β^--decay occurs, a down quark emits a field particle called a W^--boson (a weak force carrying particle) and transforms into an up quark. The W^--boson quickly breaks up into an electron and an antineutrino. The process is shown in Figure.28.4.

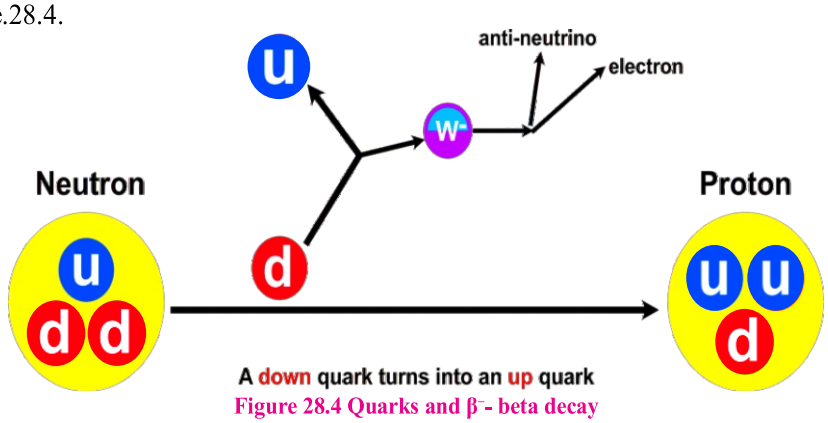


Figure 28.4 Quarks and β^-- beta decay

Electric charge is conserved during beta decay with the emission of an electron and an antineutrino.

β^+- decay is a type of radioactive decay process in which a proton is transformed into a neutron. In β^+-decay, an up quark emits a field particle called a W^+-boson and transforms into a down quark. The W^+-boson quickly breaks up into an anti-electron (positron) and a neutrino. The process is shown in Figure.28.4

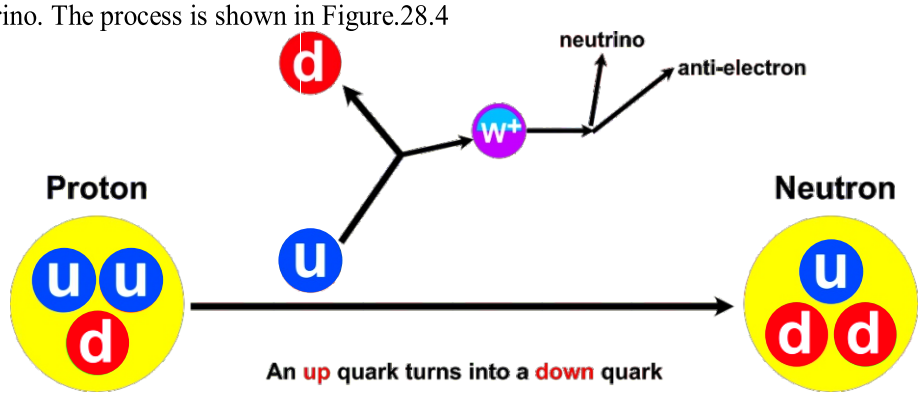


Figure28.5 Quarks and β^+- beta decay

The up quark can change into a down quark by emitting W+ weak particle. How could that trick be done? Explain.

Nature can create short-lived particles, like particle-antiparticle pairs, out of nothing by borrowing energy permitted by the uncertainty principle. This principle allows for a fluctuation in energy levels at any given moment, making it possible for particles to briefly come into existence before annihilating each other.

Imagine that there was created such a pair of particles: a down quark and anti-down quark in the neighborhood of the up quark:

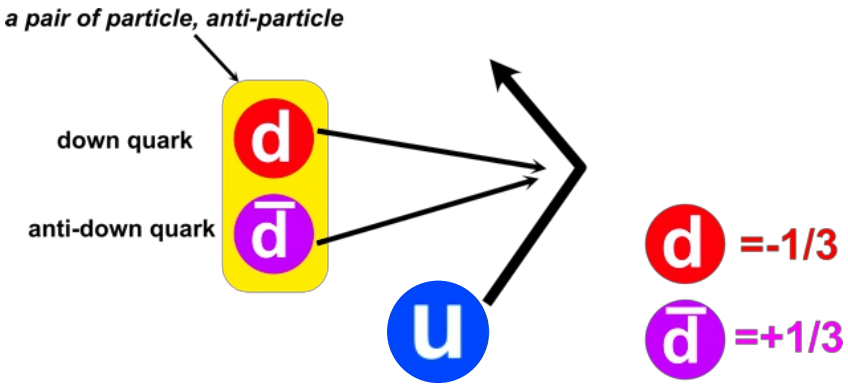


Figure 28.6 A pair of particles and anti-particles

Anti-down quark is represented by bar over symbol d: $\bar{d}$. The charge on anti-down is +1/3. Therefore, net charge on the pair is zero. Next, imagine the down quark replacing the up quark. The up quark could then join up with the remaining anti-down quark to make the W^+ particle:

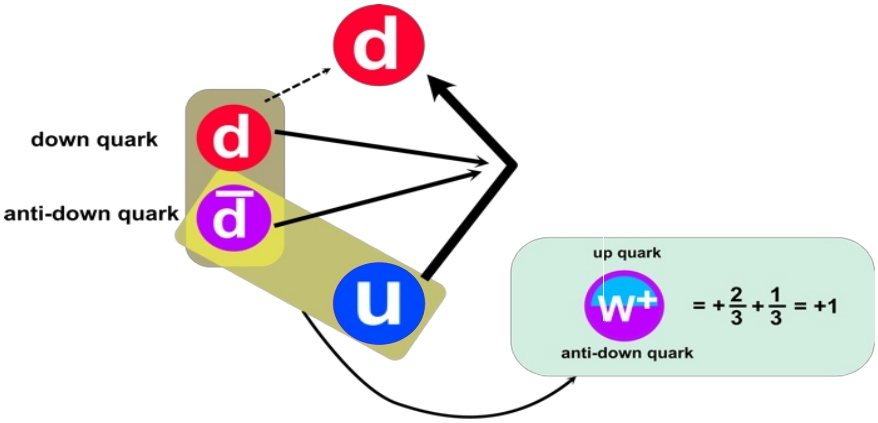


Figure 28.7 up quark and anti-down quarks

The net charge on W^+ is +1. The W^+ boson can vary its constituents and it can consist of a pair of positron-neutrino. The net charge is again +1. Now the W^+ boson decays into positron and neutrino as shown in diagram.

28.2 Radiations Detectors:

Radiations Detectors or Particle Detectors are devices that detect, track, and identify ionizing particles produced by nuclear decay, cosmic radiation, or particle accelerator reactions. In addition to reporting the presence of a particle, detectors can measure its energy and other properties such as momentum, spin, charge, and identify particle type. Particle detectors are classified into numerous types, including ionization detectors, scintillation detectors, Cherenkov light detectors, transition radiation detectors, and others.

Many particle detectors work by measuring the ionization produced when charged particles pass through a medium. The detectors based on the loss of energy caused by the ionization of atoms are called **gaseous detectors**, such as the **Wilson cloud chamber** and the **GM counter**.

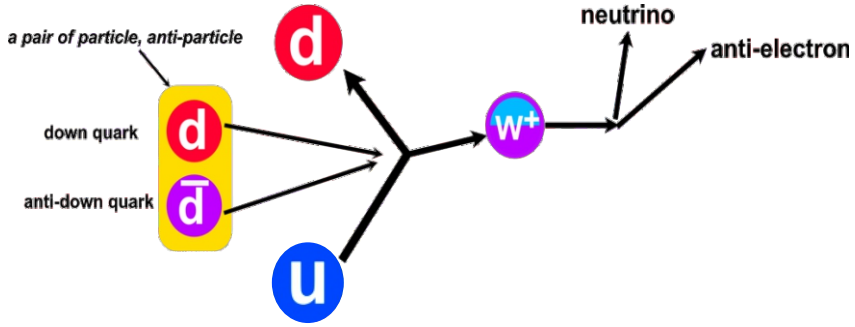


Figure 28.8 Pair of Positron and Neutrino

Gaseous detectors are a crucial tool in detecting and analyzing electronic signals. They convert the ionization produced by a charged particle through a gas into an electronic signal, providing accurate measurements of a particle's position or trajectory, therefore, they are also known as **tracking detectors**. Wilson cloud chamber is a type of tracking detector.

28.2.1 Wilson cloud chambers:

A Wilson cloud chamber is a type of tracking detector which works on the principle of ionization. It is a class of gas detector that is used in particle physics and nuclear physics to visualize the tracks of subatomic particles such as electrons, positrons, alpha particles, and cosmic rays. It was invented by Scottish physicist Charles Thomson Rees Wilsonin1912 and played a crucial role in the early discoveries of subatomic particles.

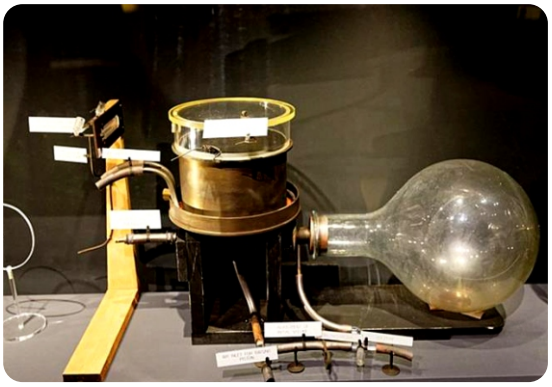


Figure 28.9 Wilson cloud Chambers

Working Principle:

The Wilson Cloud Chamber consists of a sealed container filled with a supersaturated vapor, typically water or ethanol. When a particle passes through the chamber, it ionizes the vapor, creating a trail of droplets that condense around the ionized path. This creates a visible cloud-like track that can be photographed and analyzed.

Construction:

The schematic diagram of Wilson cloud chamber is shown in figure 28.10, which consists of a large cylindrical chamber A, with walls and a ceiling made of glass. It contains dust-free air saturated with water vapor. P is a piston working inside the chamber. When the piston moves down rapidly, adiabatic expansion of the air inside the chamber takes place. The piston is connected to a large evacuated vessel F through a valve V. When the valve is opened, the air under the piston rushes into the evacuated vessel F, thereby causing the piston to drop suddenly. The wooden blocks WW reduce the air space inside the piston. Water at the bottom of the apparatus ensures saturation in the chamber. The expansion ratio can be adjusted by altering the height of the piston.

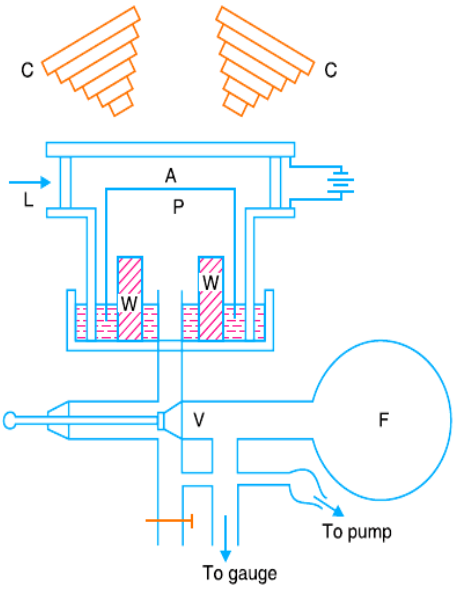


Figure 28.10 Schematic diagram of Wilson Cloud Chamber

As soon as the gas in the expansion chamber is subjected to sudden expansion, the ionizing particles are shot into the chamber through a side window. A large number of extremely fine droplets are formed on all the ions produced by the ionizing particles. These droplets form a track of the moving ionizing particles. At this stage, the expansion chamber is profusely illuminated by a powerful beam of light L and two cameras CC are used to photograph the tracks as shown in figure 28.10. The process of expansion, shooting of the

ionizing particles into the expansion chamber, illuminating the chamber and clicking the camera must all be carried out in rapid succession in order to get satisfactory results. The type of ionizing particle can be identified by its track in the cloud chamber. Alpha particles, being relatively massive, travel in straight, thick, and clearly defined paths. Beta particles, being lighter, are easily deflected and create thin, curved paths. The cloud chamber has been instrumental in discovering many elementary particles, such as the positron and meson.

Use of Wilson Cloud Chambers:

- **Particle Identification:** Wilson cloud chambers were historically instrumental in the identification and study of subatomic particles. By observing the curvature of particle tracks in a magnetic field and the nature of the tracks themselves, scientists could identify and classify various particles.
- **Nuclear Physics Research:** Cloud chambers have been used to study the behavior of particles in nuclear reactions and to investigate the structure of atomic nuclei.
- **Cosmic Ray Studies:** Wilson cloud chambers are also used in cosmic ray research. These instruments can detect and track the passage of cosmic rays, which are high-energy particles originating from space.
- **Education and Outreach:** Cloud chambers are often used as educational tools in physics classrooms and science museums to help students and the general public visualize the behavior of subatomic particles.

Geiger-Muller counter:

A Geiger-Muller counter (GM counter) is a type of gas radiation detector used to measure ionizing radiation. It is a portable and versatile device commonly employed for detecting the presence and intensity of various types of ionizing radiation, including alpha particles, beta particles, and gamma rays. The Geiger-Muller (GM) counter was invented by two German scientists, Hans Geiger and Walther Muller, in 1928.

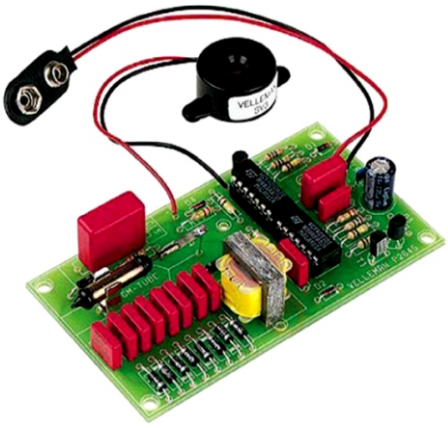


Figure 28.11 GM Counter

Construction:

The GM counter consists of a hollow metallic chamber as shown in the figure 28.11 that acts as a cathode.
A thin wire anode is also placed along its axis.
The chamber has a sealed window, through which the radiation enters the chamber.
The chamber is filled with an inert gas at low pressure.
There is a counter connected to this system to measure the radiation.

Working:

The chamber is filled with an inert gas (helium, neon, or argon) at low pressure. A high voltage is applied to this chamber. The metallic chamber will conduct electricity. When radiation enters the chamber through the window, the photons in the radiation will ionize the inert gas inside the chamber. This will make the gas conductive. The electrons produced due to ionization are accelerated due to the potential that we applied and these electrons cause

even more ionization. The ionized electrons travel towards the anode. The anode is connected to a counter. The counter counts the electrons reaching the anode. This is how we measure radiation.

Resultant Curve: Geiger Plateau

The performance of a GM counter is typically illustrated by the Geiger Plateau curve, which shows the count rate versus the applied voltage.

1. **Initial Region:**
   At low voltages, no significant ionization occurs, so the count rate is very low.
2. **Threshold Region:**
   As voltage increases, the counter starts to detect ionizing events. The count rate begins to rise sharply.
3. **Geiger Plateau:**
   At a certain voltage range, the count rate levels off and remains relatively constant. This is the operational region of the GM counter. The plateau indicates that each ionizing event is creating a detectable pulse, and the counter is functioning reliably.
4. **Continuous Discharge Region:**
   If the voltage is increased too much, continuous discharge occurs, where the counter no longer operates correctly, and the count rate rises rapidly again.

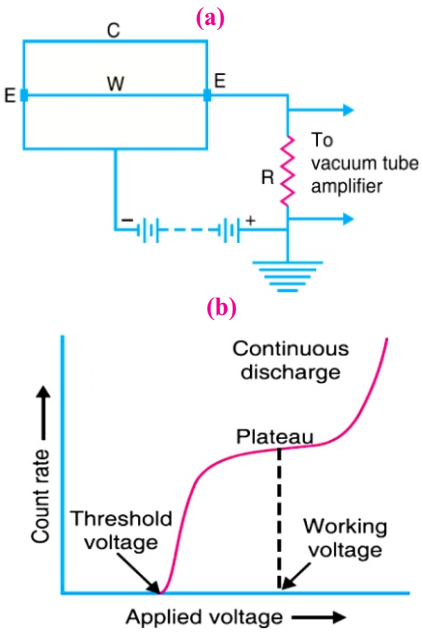


Figure 28.12 (a, b)
applied voltage vs counter rate graph

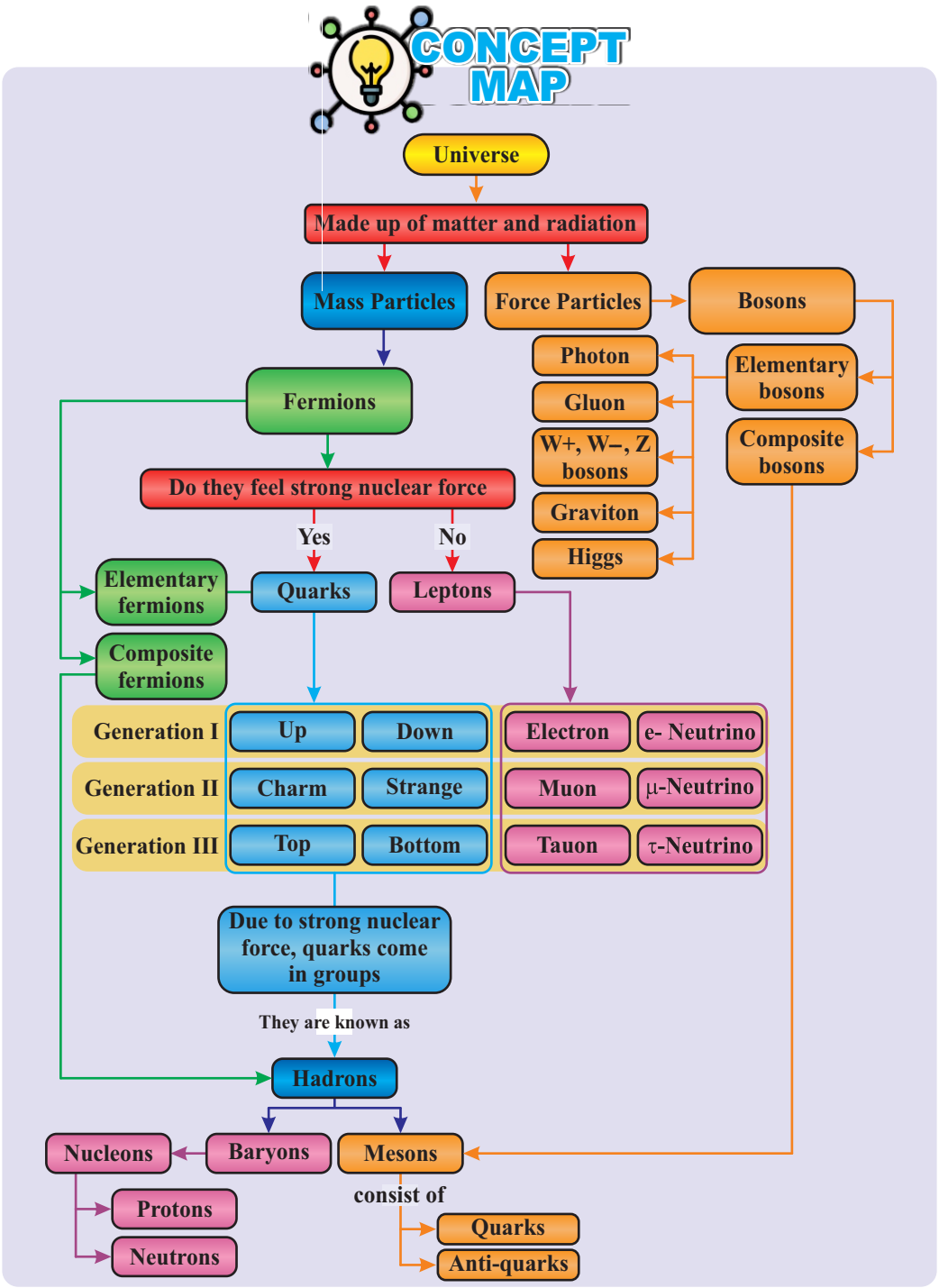
Use of GM Counter:

Geiger-Muller counters have a wide range of applications, including:

- **Radiation Monitoring:**
  They are commonly used for radiation monitoring in nuclear power plants, laboratories, and industrial settings to measure radiation levels and ensure the safety of workers and the environment."
- **Environmental Monitoring:**
  GM counters are employed for environmental radiation monitoring to assess background radiation levels and detect any abnormal increases in radiation.
- **Health Physics:**
  Health physicists use GM counters to monitor radiation exposure of individuals working with radioactive materials or in radiation-prone environments.
- **Education:**
  Health physicists use GM counters to monitor radiation exposure of individuals working with radioactive materials or in radiation-prone environments.
- **Radiological Emergencies:**
  In the event of radiological emergencies or accidents involving radioactive materials, GM counters can be used to assess radiation contamination levels.



- ✓ Gravitational Force it is weakest, acts on mass, mediated by the hypothetical graviton.
- ✓ Electromagnetic it is a force acts on electric charge, mediated by the photon.
- ✓ Strong Nuclear Force is a strongest force; it acts on quarks and gluons within hadrons mediated by gluons.
- ✓ Weak Nuclear Force is responsible for radioactive decay, mediated by W and Z bosons.
- ✓ Standard Model of Matter Hadrons is composite particles made up of quarks, including baryons (e.g., protons and neutrons) and mesons.
- ✓ Leptons are fundamental particles are not made of quarks, including electrons, muons, taus, and their associated neutrinos.
- ✓ Quarks are fundamental particles that combine to form hadrons, come in six flavors (up, down, charm, strange, top, bottom).
- ✓ Radiation Detectors is devices used to detect and measure ionizing radiation for safety, environmental monitoring, and scientific research.
- ✓ Wilson Cloud Chambers
  - Working Principle: It visualizes paths of charged particles through supersaturated vapor, forming condensation trails.
  - Construction: Sealed chamber with alcohol vapor and cooled base.
  - Use: Detecting and studying the tracks of ionizing particles, observing particle interactions.
- ✓ Geiger-Müller (GM) Counter
  - Working Principle: Detects ionizing radiation by ionizing gas within a tube, generating electrical pulses.
  - Construction: Tube filled with inert gas, central wire anode, outer cathode, high voltage applied.
  - Use: Measuring radiation levels, monitoring radioactive contamination, ensuring safety in nuclear facilities and laboratories.


CONCEPT MAP
Universe
Made up of matter and radiation
Mass Particles
Force Particles
Bosons
Photon
Gluon
W+, W–, Z bosons
Graviton
Higgs
Elementary bosons
Composite bosons
Fermions
Do they feel strong nuclear force
Yes
No
Elementary fermions
Quarks
Leptons
Composite fermions
Generation I
Up
Down
Electron
e- Neutrino
Generation II
Charm
Strange
Muon
μ-Neutrino
Generation III
Top
Bottom
Tauon
τ-Neutrino
Due to strong nuclear force, quarks come in groups
They are known as
Hadrons
Nucleons
Baryons
Mesons
Protons
Neutrons
consist of
Quarks
Anti-quarks



Section (A): **Multiple Choice Questions (MCQs)**
Choose the correct answer:

1. The Standard Model classifies elementary particles into two main groups:
   (a) Baryons and Leptons (b) Fermions and Bosons
   (c) Quarks and Gluons (d) Hadrons and Mesons
2. The following is NOT a flavor of quark:
   (a) Up (b) Down (c) Electron (d) Top
3. The charge of an up quark is:
   (a) +1/2e (b) -1/2 e (c) +2/3e (d) -2/3e
4. The concept of "quark confinement" implies that:
   (a) Quarks cannot exist as free particles outside of hadrons.
   (b) Quarks are always found in pairs with opposite charges.
   (c) Quarks have a strong affinity for gluons.
   (d) Quarks are the fundamental building blocks of all matter.
5. The primary role of the Higgs boson in the Standard Model is:
   (a) Mediating electromagnetic interactions (b) Providing mass to other particles
   (c) Transmitting the strong nuclear force (d) Creating dark matter
6. According to the Standard Model, the term "color" refers to:
   (a) Visible light spectrum (b) Charge property of quarks
   (c) Mass of particles (d) Spin of particles
7. A particle made up of a quark and an antiquark is called:
   (a) Lepton (b) Baryon
   (c) Meson (d) Neutrino
8. An elementary particle that feels all three fundamental forces (electromagnetic, weak, and strong nuclear forces) is:
   (a) Lepton (b) Quark
   (c) Electron (d) Neutrino
9. The primary function of a Geiger-Müller counter in particle physics is to:
   (a) Measure the velocity of particles
   (b) Detect and count ionizing radiation
   (c) Create antimatter particles
   (d) Generate magnetic fields
10. In a Geiger-Müller counter, the gas commonly used to detect ionizing radiation is:
    (a) Oxygen (b) Neon (c) Argon (d) Helium

Section (B): CRQs (Short Answered Questions):

1. Explain the difference between bosons and fermions and their roles in mediating fundamental forces.
2. Compare and contrast the properties of quarks and leptons, the two main categories of fermions in the Standard Model.
3. Define the term "lepton" and provide examples of leptons. Explain their fundamental properties and role in the Standard Model of particle physics.
4. Explain the concept of color charge in quarks and its significance in the strong nuclear force. How does the combination of quarks contribute to the color-neutral nature of protons and neutrons?
5. Describe the structure of a proton and neutron in terms of its quark composition. How do quarks combine to form a proton and a neutron, and what are the specific types of quarks involved?

Section (C): ERQs (Long Answered Questions):

1. Explain the structure of the Standard Model, including the different types of particles and their relationships. How does the model classify fundamental forces?
2. Describe in detail all fundamental forces and their associated field particles. What is a boson, and why are bosons referred to as field particles?
3. Describe the operating principle of a Geiger-Müller counter. How does it detect and quantify ionizing radiation, and what are its limitations in terms of measurement range and types of radiation detected?
4. Discuss the uses of a Wilson cloud chamber in particle physics experiments. How does it visualize charged particle tracks, and what factors can affect its performance?