

# SCHOOL OF SCIENCE & HUMANITIES DEPARTMENT OF PHYSICS

**UNIT-I - Allied Physics I - SPH1217** 

# **PROPERTIES OF MATTER**

# **ELASTICITY:**

**Elasticity** is the property of materials to regain its original condition (of length, volume and shape) on the removal of deforming forces.

On the removal of deforming forces, some materials completely recover their original condition called **perfectly elastic materials** and some materials do not show any tendency to recover their original condition called **plastic materials**.

**Deforming force (Load)** is a combination of all external forces acting on the material changing the dimension.

# Stress (σ):

It is the deforming force applied per unit area of the material or is the restoring force per unit area set up inside the deformed material.

 $\sigma = (F/A)$ 

Unit is  $Nm^{-2}$  or Pascal.

# Strain (ε):

It is the ratio of change in dimension to original dimension of a body when it is deformed. It is a dimensionless quantity and no unit.

# **HOOKE'S LAW:**

Hooke's law gives a relationship between the stress and strain.

According to Hooke's law, within the elastic limit, strain produced in a body is directly proportional to the stress produced.

i.e. Stress  $\alpha$  Strain

(or) (Stress/Strain) = constant, known as Modulus of Elasticity. Its unit is Nm<sup>-2</sup>.



# Types of modulus of elasticity:

Depending upon the three types of strains (linear, volume and shear), modulus of elasticity is classified into 3 categories. They are as follows:

- 1. Young's Modulus (E or y)
- 2. Bulk Modulus (k) and
- 3. Rigidity Modulus or Shear Modulus (n)

# **ELASTIC CONSTANTS:**

Young's modulus, Bulk modulus, and Rigidity modulus of an elastic material are together called as Elastic constants.

# Young's Modulus (E or y):

It is a measure of stiffness of an isotropic elastic material. Within the elastic limit, it is defined as the ratio of the longitudinal stress to longitudinal strain.

E = y = (longitudinal stress/ longitudinal strain) = (F/A)/(dL/L) = (FL)/(A dL)

If A=1, L=1 and dL=1 then y=F. So Young's Modulus of a material is equal to the force required to extend a material of unit length and unit area of cross section through unity. Unit is Pascal (Nm<sup>-2</sup>)



# **Bulk Modulus (k):**

It is a measure of resistance of the substance to uniform compression. It is defined as the ratio of the normal stress to volumetric strain.

k = (normal stress/ volumetric strain) = (F/A)/(dV/V) = (FV)/(A dV)

It is also referred to incompressibility and hence its reciprocal is **compressibility** (1/k).



# **Rigidity Modulus or Shear Modulus (n):**

It is concerned with the deformation of a solid when it experiences a force parallel to one of its surface while its opposite faces experience an opposing force (friction). Within the elastic limit, it is defined as the ratio of the tangential stress to shear strain.

n = (tangential stress/ shear strain) =  $(F/A)/(\theta) = (F)/(A\theta)$ 

When A=1 and  $\theta$ =1 then n=F,

Rigidity Modulus of a material is defined as the shearing stress per unit shear.



# **Poisson's ratio (γ):**

It is the ratio of the relative contractive strain or transverse strain (normal to applied load) to the relative extension strain or axial strain (in the direction of the applied load). When a material is stretched in one direction, it tends to contract in the other directions and vice versa.

Consider a wire of length L and diameter D is stretched by a force, so that the length increased by dL while diameter decreased by dD.

 $\gamma$  = (relative contractive strain / relative extension strain)

= (transverse strain / axial strain)

= (Lateral strain / Longitudinal strain) = (dD/D)/(dL/L)



# **BENDING OF BEAM:**

Beam is defined as a rod or bar made of homogeneous and isotropic elastic material with uniform cross section.

The length of a beam is very large compared to its breath and thickness.

Thus the shearing stress at any point of the beam is very small and negligible.

# **Assumptions:**

1. The weight of the beam is negligible in comparison to the load.

- 2. The beam should have small curvature.
- 3. There are no shearing forces.
- 4. The minimum deflection of the beam is small compared to its length

5. The cross section of the beam remains unaltered during bending. So moment of inertia of the beam remains constant.

Consider one end of the beam is fixed while a load is applied at the other end. So longitudinal filament is extended at the convex side while a contraction is obtained at the concave side. There is no change in the filament at the centre. The central filament is in the neutral stage called neutral filament and corresponding central axis is known as **neutral axis**.

# **BENDING MOMENT OF A BEAM:**

The beam ABCD is fixed at one end AD, while a load W is applied at the other end BC. Let EF be the neutral axis of the beam.

A D E D in tral axis load (w)

Consider PBCQ is a small portion of the beam in the bending, RF is its neutral axis,  $\Phi$  is the angle substantial by it at the centre of curvature O and R is the distance between neutral axis RF and centre of curvature O (radius of curvature of the beam).

When the beam is at equilibrium,

Bending moment = Restoring couple = Moment of couple

**Couple:** Two equal and opposite forces will constitute a couple.

Consider a thin filament XY in the upper half portion of the beam (above the neutral axis) situated at a distance r from the neutral axis RF.

Length of filament (XY) without bending = length of neutral axis (RF)

Original length of the filament =  $R\Phi$ 

Length of filament (XY) with bending =  $(R+r)\Phi$ 

Change in length of the filament =  $[(R+r)\Phi] - [R\Phi] = r$ 

Therefore, Strain produced at the filament (XY) = (Change in length/Original length) = (r/R)

WKT, Young's modulus = Y = (Stress/Strain) = (F/a)/(r/R) = (FR/ar)

Where a is the area of cross section of the filament

Force on the area (a) = F = (Yar / R)

Therefore,

Moment of this force about neutral axis  $RF = (Force) \times (Distance) = (Yar / R) \times (r) = (Yar^2/R)$ 

WKT, the moment of inertia of the force acting on the upper and lower halves of the beam is same.

Hence, the Total moment of inertia =  $\sum (Yar^2/R) = (Y/R) \sum (ar^2) = (Y/R) I_g$ 

Where,  $I_g$  is the geometrical moment of inertia of the beam (=  $AK^2$ )

A is the area of cross section of the beam

K is the radius of gyration of the beam

In equilibrium, bending moment of the beam is equal and opposite to the moment of bending couple.

Therefore, Bending moment of the beam =  $(Y/R) I_g$ 

The quantity  $(YI_g)$  is called **flexural rigidity** of the beam.

Flexural rigidity is defined as the bending moment required to produced a unit radius of curvature.

# **CANTILEVER:**

A cantilever is a beam which is fixed horizontally at one end and a load is applied at other end.

# DEPRESSION AT THE LOADED END OF A CANTILEVER:

Consider a cantilever of length L and PQ is the neutral axis of the cantilever.

When a load W is applied at the free end Q of the cantilever. Then the end Q is deflected to Q'.

Thus neutral axis takes a new position PQ'.

Consider a point U in the cantilever which is at a distance x from the fixed end P.



Moment of couple at U due to load W = Force x Distance= (W)(UQ) = W(L-x)

WKT, under equilibrium condition,

Moment of couple = Bending moment = Restoring couple

 $W(L-x) = (Y/R) I_g$  ------(1)

Y is the young's modulus of the rod of the cantilever

Ig is the geometric moment of inertia

R is the radius of curvature of the neutral axis

Consider a point V at a distance dx from U. the radius of curvature at V is same as that of at U, since the point V is very close to the point U.

Hence,  $UV = dx = R d\theta$ 

(or)  $R = (dx/d\theta)$ 

On substituting this in eqn (1) we get,

$$W(L-x) = (YI_g d\theta / dx)$$

(or)

$$d\theta = W(L-x) dx / YI_g$$

Draw tangents at U and V, which meet the vertical line QQ' at S and T respectively. So observed depression dy of V below U is

 $dy = (L-x) d\theta = W(L-x)^2 dx / YI_g$ 

Total depression of the beam QQ' is obtained by integrating the above equation with limit 0 to L.

Total depression = 
$$\int dy = y = {}_0 \int^L W(L-x)^2 dx / YI_g$$

Therefore, Total depression (y) =  $(W/YI_g)_0 \int^L (L^2 - 2Lx + x^2) dx$ 

(or)  $y = (W/YI_g) [L^2x - 2Lx^2/2 + x^3/3]_0^L$ 

. .

(or) 
$$y = (W/YI_g) [L^3 - L^3 + L^3/3]$$

(or) 
$$y = (W L^3 / 3YI_g)$$

(or)  $y = (mg L^3 / 3YI_g) (W = mg)$ 

Where, m is the mass and g is the acceleration due to gravity

For a circular cross section,  $I_g = (\Pi r^4/4)$ 

For a rectangular cross section,  $I_g = (bd^3/12)$ 

Where, r is the radius of circular cross section

b and d are the breath and thickness of rectangular cross section

The Young's modulus of the cantilever with rectangular

cross section is determined by

 $Y = (mg L^3) / 3y (bd^3/12)$ 

The breath (b) and thickness (d) of the can be determined using Vernier caliper and Screw gauge respectively.

# **Uniform bending:**

Consider a beam is placed above the two knife edges C and D. Consider that, an equal load W is applied at the two ends. The bending of the beam is uniform and forms an arc of a circle. This bending is known as **uniform bending**.

N ge IN ×

# Non-uniform bending:

Consider a beam AB is supported symmetrically on two knife edges C and D. The load W is applied at the centre of the beam. This bending is known as **non-uniform bending** since the bending of the beam does not describe an arc of a circle. A depression is produced at the centre depending on the applied load.



# **DETERMINATION OF YOUNG'S MODULUS (Y):**

# (a) By Non-uniform bending method:

Due to the application of load at the centre, the beam appears as two separate inverted cantilevers. It appears that, the cantilevers are fixed at the centre E and loaded at the ends each with a load (W/2). The length of the cantilever is (L/2).

Young's modulus of the beam under non uniform bending by W = (W/2) and L = (L/2)

WKT, the depression  $(y) = (W L^3 / 3YI_g)$ 

(or) Young's modulus (Y) = (W  $L^3/3yI_g$ )

In this case, W = (W/2) and L = (L/2)

Hence, the above equation can be written as

 $Y = (W/2) (L/2)^3 / 3yI_g$ 

(or)

$$\mathbf{Y} = (\mathbf{W} \ \mathbf{L}^3 / \mathbf{48yI_g})$$

(or)  $Y = (mg L^3 / 48yI_g)$ 

Where, m is the mass and g is the acceleration due to gravity

For a circular cross section,  $I_g = (\Pi r^4/4)$ 

For a rectangular cross section,  $I_g = (bd^3/12)$ 

Where, r is the radius of circular cross section

b and d are the breath and thickness of rectangular cross section

# (b) By Uniform bending method:

Young's modulus (Y) = (W x  $L^2/8yI_g$ )

(or)  $Y = (mg \ x \ L^2 / 8yI_g)$ 

Where, m is the mass and g is the acceleration due to gravity

For a circular cross section,  $I_g = (\Pi r^4/4)$ 

For a rectangular cross section,  $I_g = (bd^3/12)$ 

Where, r is the radius of circular cross section

b and d are the breath and thickness of rectangular cross section

# **TORSION:**

It is the application of a twisting or turning force to an object. This twisting force most often involves movement in one direction at one end of an object and stability or movement in the opposite direction at the other end.

# TORSION COUPLE (OR) TWISITNG COUPLE PER UNIT TWIST (C) :

Torsional rigidity is also known as the twisting couple per unit twist. It is the amount of twisting couple or torque required to twist the object by one unit angle, i.e., one radian.





Let us consider a cylindrical wire of length, L and radius, r. The wire is fixed at its upper end and twisted through an angle,  $\theta$  by applying a torque at the lower end. The wire can be assumed to be made up of hollow cylindrical tubes whose radii vary from 0 to r.

# The twisting couple on the wire, $C = \Pi n \theta r^4 / 2L$

If  $\theta = 1$  radian. Then

The twisting couple per unit twist on the wire,

 $C = \Pi nr^4/2L$ 

# POTENTIAL ENERGY IN A TWISTED WIRE:

The potential energy (restoring couple) in a twisted wire through an angle  $(\theta) = {}_0\int^{\theta}$  Moment of couple x d $\theta$ 

 $P.E = {}_{0} \int^{\theta} C\theta \ d\theta$  $P.E = 1/2 \ C\theta^{2}$ 

WKT,  $C = \Pi nr^4/2L$ 

Therefore, P.E =  $1/2 \Pi nr^4/2L \theta^2$ 

(or) P.E =  $\Pi nr^4 \theta^2/4L$ 

# **TORSIONAL PENDULUM:**

Consider a disc suspended from a torsion wire (thin rod) attached to its center. This setup is known as a **torsion pendulum** (or) **torsional pendulum**. A torsion wire is essentially inextensible, but is free to twist about its axis. As the wire twists it also causes the disc attached to it to rotate in the horizontal plane. Therefore, the disc is said to be executing **simple harmonic motion** (SHM). The rigidity modulus of the given wire can be determined using torsional pendulum.



# **Principle:**

When a disc (torsion pendulum) is rotated in a horizontal plane, the disc will execute a Simple Harmonic Oscillation (or) Motion due to the restoring couple produced in the wire.

# Determination of Time period of oscillation (T) and Rigidity modulus (n):

From the Law of conservation of energy,

Total energy (TE) = Kinetic energy (KE) + Potential energy (PE)

WKT, KE (Deflecting couple) =  $1/2 \text{ I}\omega^2$ 

PE (Restoring couple) =  $1/2 \ C\theta^2$ 

Therefore, TE = KE + PE

(or) TE =  $1/2 I\omega^2 + 1/2 C\theta^2$  (1)

On Differentiating Eq. (1) w.r.t. 't', we have

 $[1/2 . 2 I\omega d\omega/dt] + [1/2 . 2C\theta d\theta/dt] = 0$ 

(or)  $C\theta d\theta/dt + I\omega d\omega/dt = 0$  (2)

WKT, angular velocity,  $\omega = d\theta/dt$ 

And angular acceleration,  $d\omega/dt = d^2\theta/dt^2$ 

On substituting the value of  $\omega$  and  $d\omega/dt$  in Eq. (2), we have

 $C\theta \ d\theta/dt \ + I \ . \ d\theta/dt \ . \ d^2\theta/dt^2 = 0$ 

(or)  $d\theta/dt [C\theta + I \cdot d^2\theta/dt^2] = 0$ 

Here,  $d\theta/dt \neq 0$ . Therefore,  $C\theta + I \cdot d^2\theta/dt^2 = 0$ 

(or)  $d^2\theta/dt^2 = -C\theta/I$  \_\_\_\_(3)

WKT, the time period of oscillation (T) is given by

 $T = 2\Pi \sqrt{\text{Displacement/Acceleration}}$  (4)

On substituting Eq. (3) in Eq. (4), we have

 $T = 2\Pi \sqrt{\theta} / C\theta / I$ 

(or)  $T = 2\Pi \sqrt{I/C}$  (5)

Hence, Frequency,  $f = 1/T = 1/2\Pi \sqrt{C/I}$ 

If 'L' be the length of the wire and 'r' be the radius of the wire. Then

WKT, The twisting couple per unit twist on the wire,

 $C = \Pi n r^4 / 2L \_(6)$ 

On substituting Eq. (6) in Eq. (5), we have

 $T = 2\Pi \sqrt{2LI} / \Pi nr^{4}$ (or)  $T^{2} = 4\Pi^{2} . (2LI / \Pi nr4)$ (or)  $\mathbf{n} = (8\Pi I / r^{4}) . (L/T^{2}) N/m^{2}$ 

# VISCOUS FORCE AND VISCOSITY OF A LIQUID:

The frictional force offered by the adjacent liquid layers during its flow is called as **Viscous force** and the phenomenon is known as **Viscosity**.

**Definition:** The property of the fluid by virtue of which it opposes the relative motion between its different layers is called as **Viscosity**.

Both liquids and gases exhibit viscosity, but liquids are much more viscous than gases.

# CO-EFFICIENT OF VISCOSITY OF A LIQUID $(\eta)$ :

Consider a liquid flowing over a horizontal surface. The layer in contact with the surface is at rest. The velocities of other layers increase uniformly from layer to layer. The velocity is maximum for the top layer (Fig. 2.17).



Consider two layers of liquid separated by a distance dz (Fig. 2.18). Let v and v + dv be the velocities of two layers. So the velocity gradient is dv/dz. Let A be surface area of the layer. The viscous force is directly proportional to the surface area A and velocity gradient dv/dz.

$$F \propto A \frac{dv}{dz}$$
 or  $F = \eta A \frac{dv}{dz}$ .

Here,  $\eta$  is a constant for the liquid. It is called coefficient of viscosity. If A = 1 and dv/dz = 1, we have  $F = \eta$ .

**Definition :** The coefficient of viscosity is defined as the tangential force per unit area required to maintain a unit velocity gradient.

Unit of  $\eta$  is N s  $m^{-2}.\,$  It is called the Pascal second.

Dimensions of 
$$[\eta] = \frac{[F]}{[A][(dv/dz)]} = \frac{MLT^{-2}}{L^2(LT^{-1}/L)} = [ML^{-1}T^{-1}]$$

# COMPARISON OF VISCOSITIES OF TWO LIQUIDS BY GRADUATED BURETTE METHOD:

**Apparatus :** A graduated burette is clamped vertically. The nozzle is connected to a horizontal capillary tube *AB* by means of rubber tube (Fig. 2.22.)



The rubber tube is provided with a pinch clip. A beaker is placed below the free end B of the capillary tube.

**Experiment :** (*i*) The burette is filled with the first liquid whose viscosity is  $\eta_1$ . The pinch clip is opened completely. The liquid is allowed to flow out. The time taken for the liquid level to fall from 10 cc to 20 cc graduation is determined using stop watch. This is repeated and average time ( $t_1$  sec.) is found. Similarly, the time of flow of the liquid for the ranges 20–30, 30–40cc are also found out.

(*ii*) The burette and the capillary tube are washed, cleaned and arranged again. The pressure head is kept the same as before. The burette is then filled with the second liquid whose viscosity is  $\eta_2$ . The times of flow for the same ranges 10–20, 20–30, 30–40cc graduations are determined.

The ratio  $\frac{t_1}{t_2}$  for each range is calculated. The mean value of  $\frac{t_1}{t_2}$  is determined. The readings are tabulated as follows :

	Time for liquid I			Time for liquid II			
Range	I	2	Mean t <sub>1</sub>	1	2	Mean $t_2$	$\frac{t_1}{t_2}$
10-20							
20-30							
30-40							

Mean  $(t_1/t_2) =$ 

Let  $\rho_1$  and  $\rho_2$  be the densities of the two liquids respectively.

Using a U-tube or Hares' apparatus, the ratio of the densities  $\frac{\rho_1}{\rho_2}$  is determined.

 $\frac{\eta_1}{\eta_2}$  is calculated using the formula,

$$\frac{\eta_1}{\eta_2} = \left(\frac{t_1}{t_2}\right) \times \left(\frac{\rho_1}{\rho_2}\right).$$

**Theory :** Let l and a be the length and radius of the capillary tube respectively. Let the volume of liquid flowing in t seconds be Q. Volume of liquid flowing per second

=  $V = \frac{Q}{t} = \frac{\pi p a^4}{8\eta l}$  (Poiseuille's formula)

Pressure difference between the ends of the capillary tube is

 $p = h \rho g.$   $Q = \frac{\pi h \rho g a^4 t}{8 \eta l}$   $\eta = \frac{\pi h \rho g a^4 t}{8 l O}$ 

*.*..

For the first liquid, 
$$\eta_1 = \frac{\pi h \rho_1 g a^4 t_1}{8 l O}$$
 ...(1)

For the second liquid, 
$$\eta_2 = \frac{\pi h \rho_2 g a^3 t_2}{8 l Q}$$
 ...(2)

Dividing (1) by (2), we get

$$\frac{\eta_1}{\eta_2} = \left(\frac{t_1}{t_2}\right) \times \left(\frac{\rho_1}{\rho_2}\right).$$

# SURFACE TENSION:

Surface tension is that property of liquids owing to which they tend to acquire minimum surface area.

Small liquid drops acquire spherical shape due to surface tension. Big liquid drops flatten out due to weight.

The following experiment illustrates the tendency of a liquid to decrease its surface area.

When a camel hair brush is dipped into water, the bristles spread out [Fig. 2.24 (*a*)]. When the brush is taken out, the bristles cling together on account of the films of water between them contracting [Fig. 2.24 (*b*)]. This experiment clearly shows that the surface of a liquid behaves like an elastic membrane under tension with a tendency to contract. This tension or pull in the surface of a liquid is called its *surface tension*.





**Definition :** Surface tension is defined as the force per unit length of a line drawn in the liquid surface, acting perpendicular to it at every point and tending to pull the surface apart along the line.

**Unit of Surface Tension :** Surface tension is force per unit length. So its SI unit is newton per metre  $(Nm^{-1})$ .

**Dimensions of Surface Tension : Surface tension** is the ratio of a force to a length.

Surface tension = force/length

Dimensions of force = 
$$MLT^2$$
.

Dimensions of length 
$$= L$$

 $\therefore$  Dimensions of surface tension =  $\frac{MLT^{-2}}{L} = MT^{-2}$ .

The dimensional formula for surface tension is  $[MT^{-2}]$ .

# **INTERFACIAL SURFACE TENSION:**

In general, surface tension is the property of the liquid in contact with gas phase (usually air). Interfacial surface tension, on the other hand, is the property between any two substances. It could be liquid-liquid, liquid-solid or solid-air.

Surface tension can also be defined as the amount of energy required to increase the surface area between a liquid and a gas (e.g. air and water).

$$\sigma = \frac{mg}{3.8 r}$$

 $\sigma =$ Surface tension of the given liquid

- m = Mass of a single drop
- g = Acceleration due to gravity
- r = External radius of the cube

Interfacial tension is defined as the amount of energy required to increase the interfacial area between two immiscible liquids (e.g. oil and water).

$$\sigma = \frac{mg\left(1 - \frac{\rho_2}{\rho_1}\right)}{3.8 r}$$

- $\sigma =$  Interfacial tension between water and kerosene
- m = Mass of a single drop
- $\rho_1$  = Density of the heavier liquid
- $\rho_2$  = Density of the lighter liquid
- r = External radius of the tube.

# DETERMINATION OF SURFACE TENSION AND INTERFACIAL SURFACE TENSION BY THE METHOD OF DROPS (OR) DROP WEIGHT METHOD:

# (I) DETERMINATION OF SURFACE TENSION:



**Procedure :** (*i*) **S.T. of a liquid :** A short glass tube is connected to the lower end of a burette (or funnel), clamped vertically, by means of a rubber tube (Fig. 4.1). The funnel is filled with the liquid whose surface tension is to be determined. A beaker is arranged under the glass tube to collect the liquid dropping from the funnel. The stopcock is adjusted so that the liquid drops are formed slowly.

An empty beaker is taken and weighed in a balance. Now 50 drops of liquid are collected in the beaker. The beaker is again weighed. The difference between this weight and the weight of the empty beaker gives the weight of 50 drops of the liquid. From this the mass m of each drop is calculated. The external radius (r) of the tube is determined using a screw gauge or microscope.

Liquid	Mass of empty beaker $W_1$ kg	Mass of beaker + 50 drops $W_2$ kg	Mass of each $drop  m = \frac{W_2 - W_1}{50}  kg$	$\sigma = \frac{mg}{3.8 r}$ Nm <sup>-1</sup>
Water				
Kerosene				

# **Observations :**

Mass of empty beaker	$= \dots W_1 \text{kg}$
Mass of beaker + 50 drops of liquid	$= \dots W_2 \text{kg}$
Mass of 50 drops	= kg
$\therefore$ Average of mass of one drop ( <i>m</i> )	= kg.

# Radius of the tube :

Lea	st count of the vernier	= m.
Dia	meter (along any one direction)	= m.
Dia	meter (in a perpendicular direction)	= m.
<i>.</i> •.	Mean radius $(r)$	= m.
÷.	Surface tension = $\frac{mg}{3.8 r}$	$= \dots Nm^{-1}$

# (II) DETERMINATION OF INTERFACIAL SURFACE TENSION:





# (*ii*) Interfacial Tension between water and kerosene

Sufficient amount of the lighter liquid (kerosene) is taken in a beaker. The weight  $w_1$  of the beaker with kerosene is determined. The heavier liquid (water) is taken in the burette (Fig. 4.2). The glass tube is fixed vertically with its end under the surface of kerosene. The flow of water is regulated so that drops of water detach themselves into kerosene one by one. After collecting 50 drops, the beaker is again weighed. Let this weight be  $w_2$ . Then  $w_2 - w_1$  gives mass of 50 drops. From this the average mass *m* of each drop is calculated.

The external radius (r) of the tube is determined using a screw gauge or microscope.

The value of  $\rho_2/\rho_1$  is determined using a Hare's apparatus (Fig. 4.3).

One limb of the apparatus is dipped into a beaker of water and the other into a beaker containing

the liquid. Air is sucked out of the apparatus through the rubber tube attached to the side tube. When the liquids have risen to suitable heights in the limbs, the pinch clip is closed. The height of water column  $(h_1)$  and the height of liquid column  $(h_2)$  from the respective levels in the beaker are measured.

Then

$$\frac{h_1}{h_2} = \frac{\rho_2}{\rho_1}$$

The experiment is repeated by drawing in liquids to different heights.

# **Observations :**

No.	Mass o + light	f beaker er liquid	Mean w <sub>1</sub> kg	Mass of + lighter + 50 c	beaker r liquid trops	Mean w <sub>2</sub> kg	Mass of each drop $m$ = $\frac{w_2 - w_1}{50}$ kg
	Trial I	Trial II		Trial I	Trial II		50

Mass of beaker + kerosene $= \dots w_1 \, \mathrm{kg}$ Mass of beaker + kerosene + 50 drops $= \dots w_2 \, \mathrm{kg}$ of water collected inside kerosene $= \dots w_2 \, \mathrm{kg}$ Mass of 1 drop of water collected $= \dots m \, \mathrm{kg}$ Acceleration due to gravity (g) $= 9.8 \, \mathrm{ms}^{-2}$ Mean radius of the tube (r) $= \dots m \, \mathrm{kg}$ 

# Hare's apparatus : Ratio of densities of the two liquids

	Heigh	h o	
Trial No	Water	Liquid	$\frac{h_1}{h_2} = \frac{\rho_2}{\rho_1}$
1.			
2.			
3. 4.			

Mean =

# Calculation :

Interfacial surface tension = 
$$\frac{mg}{3.8 r} \left(1 - \frac{\rho_2}{\rho_1}\right) \text{Nm}^{-1}$$



# SCHOOL OF SCIENCE & HUMANITIES DEPARTMENT OF PHYSICS

UNIT - II - Allied Physics I - SPH1217

# HEAT

# HEAT:

It is the energy that is transferred from one body to another as the result of a difference in temperature.

If two bodies at different temperatures are brought together, energy is transferred i.e., heat flows from the hotter body to the colder.

The effect of this transfer of energy results in an increase in the temperature of the colder body and a decrease in the temperature of the hotter body.

The important distinction between heat and temperature (heat being a form of energy and temperature a measure of the amount of that energy present in a body).

# **SPECIFIC HEAT (c):**

The specific heat is the amount of heat per unit mass required to raise the temperature by one degree Celsius.

The relationship between heat and temperature change is usually expressed as

# $\mathbf{Q} = \mathbf{cm}\Delta\mathbf{T}$

where Q is the heat added c is the specific heat m is the mass  $\Delta T$  is the change in temperature The SI unit for specific heat is joule per kelvin per kilogram (J/K/kg, J/(kg K), J K<sup>-1</sup> kg<sup>-1</sup>).

# Specific heat capacity (Cp):

The specific heat capacity of a substance is the heat capacity of a sample of the substance divided by the mass of the sample.

Informally, it is the amount of energy that must be added in the form of heat to one unit of mass of the substance in order to cause an increase of one unit in its temperature. The SI unit of specific heat is joule per kelvin and kilogram, J/(K kg).

# CALLENDER'S BARNE'S METHOD (OR) CONTINUOUS FLOW METHOD TO DETERMINE THE SPECIFIC HEAT OF A LIQUID (C):

The method is due to Callendar and Barnes (1899).

Liquid is passed through the continuous flow calorimeter (Fig. 13.10) at a constant rate until all conditions are steady. At this stage the temperatures  $\theta_X$  and  $\overline{\theta}_Y$  at X and Y, and the mass  $m_1$  of liquid flowing through the calorimeter in time t are measured, together with the current  $I_1$  through the heating coil and the PD  $V_1$  across it. Under steady conditions none of the electrical energy which is being supplied is being used to heat the calorimeter, and therefore

$$V_1 I_1 t = m_1 c \left(\theta_Y - \theta_X\right) + Q \qquad [13.10]$$

where Q is the heat lost to the surroundings in time t.

**Note:** WKT, Energy = Power x Time and Power = Voltage x Current Therefore, Energy = Voltage x Current x Time



The rate of flow is altered so that the mass of liquid flowing in time t is  $m_2$ . The current and PD are adjusted (to  $I_2$  and  $V_2$ ) to bring the temperature at Y back to its original value  $\theta_Y$ . The temperature at X is that of the tank supplying the liquid and is constant at  $\theta_X$ . Since all temperatures are the same as they were with the initial flow rate, the heat lost in time t is again Q. Therefore

$$V_2 I_2 t = m_2 c \left(\theta_{\mathrm{Y}} - \theta_{\mathrm{X}}\right) + Q \qquad [13.11]$$

Subtracting equation [13.10] from equation [13.11] gives

$$(V_2 I_2 - V_1 I_1)t = (m_2 - m_1)c(\theta_Y - \theta_X)$$

Hence c.

# Advantages

- (i) The presence of the vacuum prevents heat losses by convection, and the effect of losses due to conduction and radiation is eliminated.
- (ii) The temperatures which are measured are <u>steady</u> and therefore can be determined accurately by using platinum resistance thermometers. This allows <u>small</u> temperature rises to be used (typically 2 °C) and the method is therefore suitable for determining the manner in which the specific heat capacity changes with temperature.
- (iii) The calculation does not involve the heat capacities of the various parts of the apparatus and so there is no need to know their values.

# Disadvantage

A large quantity of liquid is required.

# Further Points

- The percentage error is least when the difference between the two flow-rates is large.
- (ii) Continuous flow methods can also be used for gases.

# **NEWTON'S LAW OF COOLING:**

This applies when a body is <u>cooling</u> under conditions <u>of</u> forced convection (i.e. when it is in a steady draught). It states that the rate <u>of</u> loss <u>of</u> heat <u>of</u> a body is proportional to the difference in temperature between the body and its surroundings, i.e.

$$\begin{pmatrix} \text{Rate of loss of} \\ \text{heat to surroundings} \end{pmatrix} \propto \begin{pmatrix} \text{Excess} \\ \text{temperature} \end{pmatrix}$$

or

$$\begin{pmatrix} \text{Rate of loss of} \\ \text{heat to surroundings} \end{pmatrix} = k(\theta - \theta_0)$$

where

 $\theta = \text{temperature of body}$ 

 $\theta_0 = \text{temperature of surroundings}$ 

k = a constant of proportionality whose value depends on both the nature and the area of the body's surface.

In Other words,

This law, that rate of cooling is proportional to the temperature difference between the body and its surroundings, or  $\frac{dQ}{dt} = K(\theta - \theta_o)$ .

DETERMINATION OF SPECIFIC HEAT OF A LIQUID (S) USING NEWTON'S LAW OF COOLING:



#### Experiment: Determination of Specific heat of liquids by Newton's law of cooling

**Theory:** The amount of heat, needed to increase the temperature of unit mass of a material by 1°C is called the specific heat of that material. It is denoted by S.

In the same environment, rate of change of cooling of an object is directly proportional to the difference of temperature between the object and the surrounding. This is the theory of cooling method. Difference of temperature of the object and the surrounding must be small. When a liquid is heated of higher temperature and placed to cool. Then the rate of heat lost by a temperature of the liquid is directly proportional to the difference in temperature of the surrounding.

Suppose, water equivalent of the calorimeter along with the stirrer = W kg

Mass of the experimental liquid in the calorimeter = M kg

# Specific heat of the liquid = S J kg<sup>-1</sup>K<sup>-1</sup>

Time taken to cool the liquid from temperature-  $\theta_1^0$  to  $\theta_2^0 = t_1$  sec

Mass of water having volume equal to that of the liquid = m1 kg

## Specific heat of water = S<sub>1</sub> J kg<sup>-1</sup> K<sup>-1</sup>

Time taken to cool water from temperatures,  $\theta_1^0$  to  $\theta_2^0 = t_2$  sec

# **Note:** Rate of cooling = $dQ/dt = MS\theta / t$

So, rate of cooling of the liquid =  $[(MS + W) (\theta_1 - \theta_2)] / t_1 Js^{-1}$ 

and rate of cooling of water =  $[(M_1S_1 + W)(\theta_1 - \theta_2)]/t_2 Js^{-1}$ 

According Newton's law of cooling, rate of cooling in these two cases is equal.

so, 
$$[(MS + W) (\theta_1 - \theta_2)] / t_1 Js^{-1} = [(M_1S_1 + W) (\theta_1 - \theta_2)] / t_2 Js^{-1}$$

or, 
$$[(MS + W) (\theta_1 - \theta_2)] / t_1 = [(M_1S_1 + W) (\theta_1 - \theta_2)] / t_2$$

or, S =  $1/M [{(m_1S_1 + W)t_1/t_2} - W]$ 

or,  $S = 1/M [t_1/t_2 (m_1S_1 + W) - W] J kg^{-1} K^{-1}...$  (1)

Now putting the values of M, m<sub>1</sub>, S<sub>1</sub>, W, t<sub>1</sub> and t<sub>2</sub> in equation (1) we can find the value of S.

**Apparatus:** (1) a calorimeter with a stirrer, (2) a chamber having two walls, (3) a sensitive thermometer, (4) balance, (5) burner, (6) stop-watch etc.



### Procedure:

(1) Weight is taken of a clean and dry calorimeter along with its stirrer.

(2) Then water is heated in another container between temperature of 70°C to 75°C and that water is poured in the calorimeter upto a fixed mark. This calorimeter is placed in a two-walled chamber.

(3) Water is stirred slowly and slowly by a stirrer and temperature of water is recorded in 1°C interval. As the temperature of water is more than room temperature, so temperature of water gradually decreases. In this way 20 to 25 readings of temperature are recorded and then weight of the calorimeter along with water is taken. From the difference of subtraction of these two readings weight of water is found out.

(4) Now water is thrown away from the calorimeter and it is cleaned and dried. Then experimental heated liquid from 70°C to 75°C is poured in the calorimeter upto the previous mark. The calorimeter along with the liquid is placed inside chamber.

(5) Then the liquid is stirred slowly and following procedure (3) temperature is recorded in each degree interval of temperature. In this way 20-25 readings are taken. Afterwards calorimeter with liquid is taken. From the difference of weights of 3rd and the first one weight of the liquid is found out.

#### **Observation and Manipulation:**

Table: Mass of water and liquid

No. of	Mass of the	(Mass of the	(Mass of the	Mass of	Mass of
observation	Calorimeter	Calorimeter	Calorimeter	Water m <sub>1</sub> =	liquid M =
	with Stirrer	+ Stirrer +	+ Stirrer +	(W2 - W1) kg	(W3 - W1) kg
	$= W_1 Kg$	Water) = W2	liquid) = W3		
		Kg	Kg		
1					

Specific heat of the material of the calorimeter = S' J kg<sup>-1</sup> K<sup>-1</sup>

So, Water equivalent, W = W<sub>1</sub> x S<sub>1</sub> Kg

**Calculation:**  $S = 1/M [t_1/t_2 (m_1S_1 + W) - W] J kg^{-1} K^{-1}$ 

Result: Specific heat of the given liquid, S = ... ... J kg<sup>-1</sup> K<sup>-1</sup>

# **Emissive power:**

Emissive power is the total amount of thermal energy (radiation) emitted per unit area per unit time for all possible wavelengths. It is also called as **Emittance**.

# **Emissivity:**

Emissivity of a body at a given temperature is the ratio of the total emissive power of a body to the total emissive power of a perfectly black body at that temperature. Emissivity of surface will depend on the material of which it is composed.

**Note:** A **perfect blackbody** is one that absorbs all incoming light radiation and does not reflect any. At room temperature, such an object would appear to be perfectly black (hence the term blackbody). However, if heated to a high temperature, a blackbody will begin to glow with thermal radiation.

# **PRODUCTION OF LOW TEMPERATURE: JOULE – KELVIN (J-K) EFFECT OR JOULE-THOMSON (J-T) EFFECT:**



Joule-Thomson effect is a temperature change of a gas (usually cooling) when it is allowed to expand freely, i.e. at a gas stationary adiabatic flowing through a porous diaphragm (J. Joule, W. Thomson, 1852-1862).

During this process, the gas that initially occupied the volume  $V_1$  at the pressure  $p_1$  flows through a porous diaphragm and after that has the volume  $V_2$  at the pressure  $p_2$ . The work  $p_1V_1 - p_2V_2$  is done on the system and it equals to the change of the gas internal energy  $U_1 - U_2$  since the porous diaphragm suppresses all macroscopic motions. Therefore when gas flows in the conditions of thermal isolation, the *Enthalpy* remains constant: H=U+pV. From the condition of constancy H it follows that the change of temperature T for a unit pressure (the differential Joule-Thomson effect for the small values of  $\Delta T$  and  $\Delta p$ ) is:  $(\frac{\Delta T}{\Delta p})_H = -C_p^{-1}(\frac{\partial H}{\partial p})_T - C_p^{-1}[T(\frac{\partial V}{\partial T})_p - V]$ , where  $C_p = (\frac{\partial H}{\partial T})_p$  is the *Heat capacity* at a constant pressure. For an *Ideal gas* the Joule-Thomson effect is zero, and for the *Real gases* 

its sign depends on the sign of the expression  $[T(\frac{\partial V}{\partial T})_p - V]$ , which is defined by the

Equation of state. If the gas flow decreases the temperature, then  $\left(\frac{\Delta T}{\Delta p}\right)_H > 0$  and the Joule-

Thomson effect is positive. If the temperature rises,  $(\frac{\Delta T}{\Delta p})_H < 0$  and the Joule-Thomson

effect is negative. The temperature at which Joule-Thomson effect changes its sign is called inversion temperature and the sum of the inversion points in *p*-*T* diagram is called the inversion curve. For a gas which follows van der Waals equation, the Joule-Thomson effect is positive if  $2a(V-b)^2 > RTbV^2$  (*R* is the *Gas constant*), i.e. the constants *a* and *b* of *Van der Waals equation* produce the inverse effect on the sign of Joule-Thomson effect, defined by the competition between the molecular forces of attraction and repulsion. The inversion curve for a van der Waals gas follows the equation  $2a(V-b)^2 = RTbV^2$ . Joule-Thomson effect is one of the methods of obtaining the *Low temperatures*. To achieve the goal preferably apply Joule-Thomson effect in combination with adiabatic expansion of the gas. The differential Joule-Thomson effect is small (for the air  $\frac{\Delta T}{\Delta p} \approx 0.25 \cdot 10^{-5}$  K/Pa). Since the

integral Joule-Thomson effect pressure varies widely, it is broadly used in engineering [2].

# **Conclusions:**

At ordinary temperatures, most of the gases will cool down in passing from higher pressure to lower pressure through porous barrier. This is called as **Positive effect.** But few gases like hydrogen and helium at room temperature become hot instead of cooling. This is called as **Negative effect.** 

The temperature at which J-T effect (or) J-K effect changes its sign is called as **Inversion** temperature (or) Temperature of inversion  $(T_i)$ . It is represented as

$$T_i = 2a/Rb$$

Where a and b are the van der Waal's constants

R is the Gas constant

### LIQUEFACTION OF AIR - LINDE'S METHOD:

In this process, Joule-Kelvin effect is applied. Linde was able to liquefy air in 1896 using this effect. The apparatus used is shown in Fig. 6.2.



Fig. 6.2

The pump  $P_1$  compresses air to a pressure of about 25 atmospheres and is passed through a tube surrounded by a jacket through which cold water is circulated. This compressed air is passed through KOH solution to remove CO<sub>2</sub> and water vapour.

This air, free from  $CO_2$  and water vapour, is compressed to a pressure of 200 atmospheres by the pump  $P_2$ . This air passes through a spiral tube surrounded by a jacket containing a freezing mixture. This cooled air at high pressure and at a temperature of  $-20^{\circ}C$  is allowed to come out of the nozzle.  $N_1$ . Joule–Kelvin effect takes place and the incoming air is cooled. The cooled air is circulated to the pump  $P_2$  and is compressed. It passes through the nozzle  $N_1$  and is further cooled. This cooled air is allowed to pass through the nozzle  $N_2$  (from high pressure to low pressure side) and is further cooled. As the process continues, after a few cycles, air gets cooled to a sufficiently low temperature and after coming out of the nozzle  $N_2$  gets liquefied and is collected in the Dewar flask. The unliquefied air is again circulated back to the pump  $P_1$  and the process is repeated.

The whole of the apparatus is packed in cotton wool to avoid any conduction or radiation.

# **HELIUM I AND II:**

Liquid Helium above a temperature of 2.19 K is called Helium I. Liquid He I behaves in a normal way. Below 2.19 K liquid Helium is known as He II. He II has abnormal properties



The specific heat of liquid helium increases up to 2.19 K and *at this temperature there is a sudden* and abnormal increase in the specific heat. Beyond 2.19 K, the specific heat first decreases and then increases Fig. 7.12. (b)

The specific heat-temperature graph resembles  $\lambda$  and hence this temperature at which the specific heat ( $C_{v}$ ) changes abruptly (2.19 K) is called the  $\lambda$  **point**.

Liquid helium above 2.19 K is called helium I because it behaves in a normal way and below 2.19 K, it is called helium II because of its abnormal properties.

No heat is evolved or absorbed during the transition from Helium I to Helium II and vice-versa. In other words, no latent heat is involved during this transition. Hence the following conclusions can be drawn.

# **Conclusions:**

- (*i*) The entropy is constant across the curve *i.e.*, entropy of He II is practically the same as that of He I.
- (*ii*) The density of both the types of liquid is about the same *i.e.*, there is no change of density during transition from one form of Helium to another.

While viscosity of liquid He I decreases with decrease in temperature and this property is contrary to the property of a liquid but it resembles that of a gas. The viscosity of He II is almost zero and it can flow rapidly through narrow capillary tubes. Liquid He I is a normal liquid while He II presents a very anomalous behaviour. The thermal conductivity of He II is abnormally high, its internal friction is practically zero and that when it is forced through a capillary tube, the emerging liquid He II cools while that which remains behind warm up.

# **PRODUCTION OF LOW TEMPERATURE- ADIABATIC DEMAGNETISATION :**

Isochoric	Constant volume
Isobaric	Constant pressure
Isothermal	Constant temperature
Adiabatic	No heat transfer

Adiabatic demagnetization is a process by which the removal of a magnetic field from certain materials serves to lower their temperature. This procedure, proposed by chemists Peter Debye (1926) and William Francis Giauque (independently, 1927), provides a means for cooling an already cold material (at about 1 K) to a small fraction of 1 K.



Debye, Giauque and Macdougall were able to produce temperatures below 1 K with the help of gadolinium sulphate. Haas and Kramers used the magnetic balance for a number of paramagnetic substances. They found that potassium and chromium alum give a much lower temperature.

**Method.** The apparatus used is shown in (Fig 7.16). The paramagnetic specimen (salt) is suspended in a vessel A, which is surrounded by liquid helium. Liquid helium taken in Dewar flask D, is boiled under reduced pressure. It is surrounded by another Dewar flask  $D_2$  containing liquid hydrogen. The salt is in contact with the helium gas. A magnetic field of the order of 30,000 Gauss is applied.

When the magnetic field is switched on the specimen (salt) is magnetised. The heat due to magnetisation is removed by first introducing hydrogen gas into A and then pumping it off with high

vacuum pump, so that the specimen is thermally isolated. In the mean time, the specimen (salt) picks up the temperature of the liquid helium *i.e.*, the specimen and the liquid helium are at the same temperature. Now the magnetic field is switched off. Adiabatic demagnetisation of the specimen takes place and its temperature falls. The temperature of the specimen is determined by fitting a coaxial solenoid coil round the tube A and measuring the self-inductance and hence susceptibility of the substance at the beginning and at the end of experiment. Then temperature T is calculated by

Curie law  $\chi = \frac{C}{T}$ .

*.*...

 $\chi T$  is constant  $\frac{\chi_1}{\chi_2} = \frac{T_2}{T_1}$ 

 $\chi_1$  is the sesceptibility at temperature  $T_1$  of the helium bath.  $\chi_2$  is the susceptibility after adiabatic demagnetisation at temperature  $T_2$ .

$$\therefore \qquad T_2 = \frac{\chi_1}{\chi_2} T_1$$



# SCHOOL OF SCIENCE & HUMANITIES DEPARTMENT OF PHYSICS

UNIT - III - Allied Physics I - SPH1217

# **ELECTRICITY AND MAGNETISM**

# **ELECTRICITY:**

# **POTENTIOMETER (POT):**

**Definition:** The instrument designs for measuring the unknown voltage by comparing it with the known voltage, such type of instrument is known as the potentiometer. In other words, the potentiometer is the three terminal device used for measuring the potential differences by manually varying the resistances. The known voltage is drawn by the cell or any other supply sources.

The potentiometer uses the comparative method which is more accurate than the deflection method. So, it is mostly used in the places where higher accuracy is required or where no current flows from the source under test. The potentiometer is used in the electronic circuit, especially for controlling the volume.

# Characteristics of Potentiometer

The following are the important characteristics of the potentiometer.

1. The potentiometer is very accurate because its works on the comparing method rather than the deflection pointer method for determining the unknown voltages.

It measures the null or balance point which does not require power for the measurement.

3. The working of the potentiometer is free from the source resistance because no current flows through the potentiometer when it is balanced.

# **PRINCIPLE OF POTENTIOMETER:**

**Ans. Working principle of potentiometer** : The potential difference across a length of the potentiometer wire is directly proportional to its length (or) when a steady current is passed through a uniform wire, potential drop per unit length or potential gradient is constant.

i.e.  $\varepsilon \propto l \Rightarrow \varepsilon = \phi l$  where  $\phi$  is potential gradient.



# Comparing the emf of two cells $\epsilon_1$ and $\epsilon_2$ :

1) To compare the emf of two cells of emf  $E_1$  and  $E_2$  with potentiometer is shown in diagram



2) The points marked 1, 2, 3 form a two way key.

3) Consider first a position of the key where 1 and 3 are connected so that the galvanometer is connected to  $\varepsilon_1$ .

4) The Jockey is moved along the wire till at a point  $N_1$  at a distance  $l_1$  from A, there is no deflection in the galvanometer. Then  $\epsilon_1 \propto l_1 \Rightarrow \epsilon_1 = \phi l_1$  (1)

5) Similarly, if another emf  $\pmb{\varepsilon}_2$  is balanced against

$$l_2$$
 (AN<sub>2</sub>), then  $\varepsilon_2 \propto l_2 \Rightarrow \varepsilon_2 = \phi l_2$  (2)

6)  $\frac{(1)}{(2)} \Rightarrow \frac{\varepsilon_1}{\varepsilon_2} = \frac{l_1}{l_2}$ 

# CALIBRATION OF LOW RANGE VOLTMETER USING POTENTIOMETER:

The calibration is the process of checking the accuracy of the result by comparing it with the standard value. In other words, calibration checks the correctness of the instrument by comparing it with the reference standard. It helps us in determining the error occur in the reading and adjusts the voltages for getting the ideal reading.

# Calibration of Voltmeter

The circuit for the calibration of the voltmeter is shown in the figure below.



The circuit requires two rheostats, one for controlling the voltage and another for adjustment. The voltage ratio box is used to step-down the voltage to a suitable value. The accurate value of the voltmeter is determined by measuring the value of the voltage to the maximum possible range of the potentiometer.

The potentiometer measures the maximum possible value of voltages. The negative and positive error occurs in the readings of the voltmeter if the readings of the potentiometer and the voltmeter are not equal.

# **INTERNAL RESISTANCE:**

Internal resistance refers to the opposition to the flow of current offered by the cells and batteries themselves resulting in the generation of heat. Internal resistance is measured in Ohms. The relationship between internal resistance (r) and emf (e) of cell s given by.

e = I(r + R)

Where, e = EMF i.e. electromotive force (Volts), I = current (A), R = Load resistance, and r is the internal resistance of cell measured in ohms.

On rearranging the above equation we get;

e = IR + Ir or, e = V + Ir

In the above equation, V is the potential difference (terminal) across the cell when the current (I) is flowing through the circuit.

Note: The emf (e) of a cell is always greater than the potential difference (terminal) across the cell

# **RESISTANCE:**

**Resistance** is the property of a material by virtue of which it opposes the flow of electrons through the material. It restricts the flow of the electron through the material. It is denoted by (R) and is measured in ohms ( $\Omega$ ).

When the voltage is applied across the resistor the free electrons start accelerating. These moving electrons collide with each other and hence opposes the flow of electrons. The opposition of electrons is known as the resistance. The heat is generated when the atom or molecules are colliding with each other.

# **MEASUREMENT OF INTERNAL RESISTANCE OF CELL USING POTENTIOMETER:**

**Ans. Working principle of potentiometer :** The potential difference across a length of the potentiometer wire is directly proportional to its length (or) when a steady current is passed through a uniform wire, potential drop per unit length or potential gradient is constant.

i.e.  $\varepsilon \propto l \Rightarrow \varepsilon = \phi l$ 

where  $\phi$  is potential gradient.

Measurement of internal resistance (r) with potentiometer :

Potentiometer to measure internal resistance
 (r) of a cell (ε) is shown in diagram.

2) The cell (emf  $\varepsilon$ ) whose internal resistance (r) is to be determined is connected across a resistance box (R.B) through a key K<sub>2</sub>.

3) With key  $K_2$  open, balance is obtained at length  $l_1$  (AN<sub>1</sub>). Then  $\varepsilon = \phi l_1 \rightarrow (1)$ 

4) When key K<sub>2</sub> is closed, the cell sends a current(I) through the resitance box (R.B).

5) If V is the terminal potential difference of the cell and balance is obtained at length  $l_2$  (AN<sub>2</sub>).

Then  $V = \phi l_2 \rightarrow (2)$ 

6) 
$$\frac{(1)}{(2)} \Rightarrow \frac{\varepsilon}{V} = \frac{l_1}{l_2} \rightarrow (3)$$

7) But  $\varepsilon = I (r + R)$  and V = IR. This gives

$$\frac{\varepsilon}{\mathbf{V}} = \frac{(\mathbf{r} + \mathbf{R})}{\mathbf{R}}$$
$$\frac{l_1}{l_2} = \left(\frac{\mathbf{r}}{\mathbf{R}} + \mathbf{l}\right) [\because \text{ from (3)}]$$
$$\therefore \mathbf{r} = \mathbf{R} \left(\frac{l_1}{l_2} - \mathbf{1}\right)$$

Thus, the internal resistance (r) of the cell can be measured using potentiometer.



# **MEASUREMENT OF AN UNKNOWN RESISTANCE USING POTENTIOMETER:**

The DC potentiometer method of measurement of resistance is used for measuring the unknown resistance of low value. This can be done by comparing the unknown resistance with the standard resistance. The voltage drop across the known and unknown resistance is measured and by comparison the value of known resistance is determined.

Let understand this with the help of the circuit diagram. The R is the unknown resistance whose value is needed to be measured. The S is the standard resistance from which the value of unknown resistance is compared. The rheostat is used for controlling the magnitude of current into the circuit.



The double pole double throw switch is used in the circuit. The switch, when moves to position 1, 1 the unknown resistance connects to the circuit, and when it moves to position 2, 2 the standard resistance connects to the circuit.

Consider that when the switch is in position 1,1 the voltage drop across the unknown resistance is  $\mathsf{V}_\mathsf{r}$ 

$$V_R = IR$$

and when it is in 2, 2 the voltage drop across the resistance is  $V_{\mbox{\scriptsize s}}$ 

$$V_S = IS$$

On equating the equation (1) and (2), we get

$$\frac{V_R}{V_S} = \frac{IR}{IS}$$
$$\frac{V_R}{V_S} = \frac{R}{S}$$
$$R = \frac{V_R}{V_S} . S$$

The accuracy of unknown resistance depends on the value of standard resistance.

# **MAGNETISM:**

# **POLE STRENGTH (M):**

Pole strength of a magnet is defined as the strength with which the materials get attracted towards the magnet. It is the scalar quantity and is denoted by m. unit of pole strength will be Am.

# **MAGNETIC MOMENT (M):**

The magnetic moment (M) is the product of the pole strength (m) and the length of the magnet (l).

Therefore, the relation between the pole strength and the magnetic moment is

M = ml

Unit of magnetic dipole moment is Am<sup>2</sup>.

# **DEFLECTION MAGNETOMETER:**



Figure 2 Deflection magnetometer.

The deflection magnetometer consists of one meter long wooden board and provided at the centre of which a compass box with a small magnetic needle pivoted and provided a circular scale graduated in degrees. The circular scale is divided into quadrants with 0–0 and 90–90 readings marked at the extreme of the perpendicular diameters. The magnetic needle is free to rotate in a horizontal plane and the reading can be taken with the help of a long aluminium pointer fixed at right angles to the needle. The needle and the pointer are enclosed in a brass box with glass cover at the top. The compass box is mounted on a wooden base provided with two arms. Each arm is fitted with a half metre scale such that the zero of each scale coincides with the pivot of the magnetic needle. There is a plane mirror in the compass box which helps to take the readings vertically by making the pointer and its image fall into the line of sight (Figure 2).

Deflection magnetometer can be arranged in three positions to determine the magnetic moment (M) and the horizontal component of earth's magnetic field (BH)

(i) Tan-A position(ii) Tan-B position(iii) Tan-C position

# **TAN-A POSITION:**

![](_page_37_Figure_1.jpeg)

In Tan A position, prior to placement of the magnet, the compass box is rotated so that the (0-0) line is parallel to the arm of the magnetometer. Then the magnetometer as a whole is rotated till pointer reads (0-0). Finally, the bar magnet (the same one that was previously suspended in the Vibration Magnetometer) is placed horizontally, parallel to the arm of the deflection magnetometer, at a distance d chosen so that the deflection of the aluminium pointer is between  $30^{\circ}$  and  $60^{\circ}$ .

$$B_H \tan \theta = \frac{\mu_0}{4\pi} \frac{2md}{\left(d^2 - l^2\right)^2}$$

where l = L/2 is the half-length of the magnet m = magnetic moment of the magnet  $4\pi x \ 10^{-7} \ H/m$  - the magnetic permeability of free space

 $\theta$  = deflection of aluminium pointer.

# **TAN-B POSITION:**

![](_page_37_Figure_7.jpeg)

In this Tan B position, prior to placement of the magnet, the compass box alone is rotated so that the (90-90) line is parallel to the arm of the magnetometer. Then the magnetometer as a whole is rotated so that the pointer reads (0-0). Finally, the magnet is placed horizontally, perpendicular to the arm of the magnetometer, at distances d chosen so that the deflection of the aluminium pointer is between  $30^{\circ}$  and  $60^{\circ}$ .

$$B = B_H \tan \theta = \frac{\mu_0}{4\pi} \frac{m}{(d^2 + l^2)^{3/2}}$$

# **TAN-C POSITION:**

![](_page_38_Figure_2.jpeg)

In this Tan C position, the bar magnet is placed vertically, in contrast to the Tan A and Tan B positions, where it is placed horizontally. The bottom end of the bar magnet is placed a distance d from the center of the compass box, chosen so that the deflection of the aluminium pointer is between  $30^{\circ}$  and  $60^{\circ}$ .

$$B = \frac{\mu_o}{4\pi} P(\frac{1}{d^2} - \frac{d}{(d^2 + L^2)^{3/2}})$$

Where P is the pole strength in Amp-meters (A m) and L is the length of the bar magnet in meters

# VIBRATION MAGNETOMETER-THEORY AND PERIOD OF OSCILLATION (T):

![](_page_38_Figure_7.jpeg)

Figure 1 Box type vibration magnetometer.

The box type vibration magnetometer consists of a rectangular wooden box with glass windows and provided with levelling screws at the bottom. At the top of the wooden box, from the centre, a long brass tube is fitted vertically. A torsionless fibre is suspended from the top of the brass tube carrying a light brass stirrup H. The bar magnet can be placed horizontally on the stirrup and the oscillations can be viewed from the top glass window (Figure 1).

Thus its period of oscillation (T) for small  $\theta$ , is approximately.

$$T = 2\pi \sqrt{l \, mB_{H}} \tag{1}$$

Where I = moment of inertia of the magnet about the axis of oscillation m = magnetic moment of the magnet

BH = horizontal intensity of the earth's magnetic field.

For a rectangular bar magnet,

$$I = M \frac{L^2 + b^2}{12}$$

Where

M = mass of the magnetL = length of the magnet (longest horizontal dimension)b = breadth of the magnet (shortest horizontal dimension)

On squaring equation (1)

$$T^2 = 4\pi^2 \frac{I}{mB_H}$$

(or)

$$mB_H = 4\pi^2 \frac{I}{T^2}$$

Thus, by measuring vibration (oscillation) period *T* and calculating the moment of inertia *I* of the bar magnet,  $mB_H$  is determined using the vibration magnetometer.

# DETERMINATION OF M AND B<sub>H</sub> USING THE DEFLECTION MAGNETOMETER IN TAN C POSITION AND THE VIBRATION MAGNET:

To determine the magnetic dipole moment (M) of a bar magnet and horizontal intensity (BH) of earth's magnetic field using a deflection magnetometer.

![](_page_39_Figure_15.jpeg)

The deflection magnetometer consists of a large compass box with a small magnetic needle pivoted at the center of a circular scale so that the needle is free to rotate in a horizontal plane .

A large aluminium pointer is rigidly fixed perpendicular to the magnetic needle. The circular scale is graduated in degrees. (0-0) and (90-90) readings are marked at the ends of two perpendicular diameters. The compass box is placed at the center of a wooden board one meter long. The wooden board has a millimeter scale along its axis. The zero of this scale is at the center of the compass box.

# **TAN-C POSITION:**

![](_page_40_Figure_3.jpeg)

In this Tan C position, the bar magnet is placed vertically, in contrast to the Tan A and Tan B positions, where it is placed horizontally. The bottom end of the bar magnet is placed a distance d from the center of the compass box, chosen so that the deflection of the aluminium pointer is between  $30^{\circ}$  and  $60^{\circ}$ .

$$B = \frac{\mu_o}{4\pi} P(\frac{1}{d^2} - \frac{d}{(d^2 + L^2)^{3/2}})$$
(1)

Where P is the pole strength in Amp-meters (A m) and L is the length of the bar magnet in meters

According to the tangent law,

$$B = B_H \tan \theta \tag{2}$$

In Equation (2), the horizontal component of the field from the bar magnet  $B_H$  bar corresponds to the external field B, so we have. On substituting this in Eq. (1) and solving for the pole strength P of the bar magnet,

$$P = \frac{4\pi B_H \tan \theta}{\mu_o (\frac{1}{d^2} - \frac{d}{(d^2 + L^2)^{3/2}})}$$

By knowing the pole strength (P) of the bar magnet, the magnetic moment (M) can be determined.

![](_page_41_Picture_0.jpeg)

# SCHOOL OF SCIENCE & HUMANITIES DEPARTMENT OF PHYSICS

UNIT - IV - Allied Physics I - SPH1217

# SOUND AND ACOUSTICS OF BUILDING

# SOUND:

In physics, sound is a vibration that propagates as an acoustic wave through a transmission medium such as a gas, liquid or solid. In human physiology and psychology, sound is the reception of such waves and their perception by the brain. Only acoustic waves that have frequencies lying between about 20 Hz and 20 kHz, the audio frequency range elicit an auditory percept in humans (audible sound). Sound waves above 20 kHz are known as ultrasound and are not audible to humans. Sound waves below 20 Hz are known as infrasound. Sound can propagate through a medium as longitudinal waves and transverse waves.

# LONGITUDINAL WAVES:

**Longitudinal waves** are waves in which the displacement of the medium is in the same direction as or the opposite direction (parallel) to the direction of propagation of the wave. Mechanical longitudinal waves are also called compressional or compression waves, because they produce compression and rarefaction when travelling through a medium.

Examples: sound waves, ultrasound waves, seismic P-waves.

# **TRANSVERSE WAVES:**

**Transverse wave** is a moving wave whose oscillations are perpendicular to the direction of the wave or path of propagation. Transverse wave is also called as shear wave. In seismology, shear waves are also called secondary waves or S-waves.

**Examples:** Vibrations in a guitar string, Electromagnetic waves such as light waves, microwaves, radio waves, Seismic S-waves.

# TRANSVERSE VIBRATION OF A STRETCHED STRING:

# VELOCITY OF TRANSVERSE WAVES IN A STRETCHED STRING (v):

Consider a string subjected to a tension T. m is the mass per unit length of the string. Consider the portion PABQ of the string (Fig. 1.5).

A transverse wave is travelling in the string from left to right with a velocity v. The string is also drawn towards the left with the same velocity. So the wave is stationary.

Consider a small element AB of the string. Its length is  $\delta x$ . The centre of curvature of the element AB is O.  $\angle AOB = \theta$ . Let T be the tension at A and B. The directions of these tensions are tangential to the element at A and B.

![](_page_42_Figure_14.jpeg)

Resolve the tension at A into two rectangular components,  $T \sin \frac{\theta}{2}$  and  $T \cos \frac{\theta}{2}$  perpendicular and parallel to the string.

Similarly the components at *B* are  $T \sin \frac{\theta}{2}$  and  $T \cos \frac{\theta}{2}$  perpendicular and parallel to the string.

The parallel components are equal and opposite. The perpendicular components are along *CO*.

Resultant tension along  $CO = 2T \sin \frac{\theta}{2} = 2T \frac{\theta}{2} = T\theta$  ...(1)

This resultant tension provides the necessary centripetal force for the circular motion of the element AB.

 $\frac{m\,\delta x\,\,v^2}{R} = T\Theta$ 

m

Centripetal force =  $\frac{(m \, \delta x) \, v^2}{R}$  ...(2)

For equilibrium,

But

÷

or

$\theta = \frac{\partial x}{R}$	
$\frac{\delta x \cdot v^2}{R} = \frac{T \cdot \delta x}{R}$	
$v^2 = \frac{T}{2}$	
m	
$v = \sqrt{\frac{1}{m}}$	(3)

# **Centripetal Force:**

Centripetal force is the component of force acting on an object in curvilinear motion which is directed towards the axis of rotation or centre of curvature. The unit of centripetal force is Newton.

# **Examples of Centripetal Force:**

- > Spinning a ball on a string
- $\succ$  Turning a car
- Going through a loop on a roller
- Planets orbiting around the Sun

# **Centrifugal Force:**

Centrifugal force is a force that arises from the body's inertia and appears to act on a body that is moving in a circular path which is directed away from the centre around which the body is moving. Centrifugal force unit is Newton. The centrifugal force drives the object away from the centre. It is a fictitious force.

# **Examples of Centrifugal forces:**

- > Weight of an object at the poles and on the equator
- ➤ A bike making a turn.
- Vehicle driving around a curve
- Equatorial railway

# FREQUENCY OF TRANSVERSE VIBRATION OF A STRETCHED STRING (n):

Consider a string of length l, stretched under a tension T, fixed at its ends (Fig. 1.6).

Let *m* be the mass per unit length of the string. We pluck the string at its mid-point. The string vibrates in a *single loop* emitting its *fun-damental* note. There are nodes at the fixed ends and an antinode at the centre.

![](_page_44_Figure_3.jpeg)

 $\therefore \qquad l = \lambda/2$ Wavelength of the stationary wave  $\lambda = 2l$ 

....

$$v = n\lambda$$
 and  $v = \sqrt{\frac{T}{n}}$   
 $n = \frac{v}{\lambda}$ .  
 $n = \frac{1}{2I}\sqrt{\frac{T}{m}}$ 

# LAWS OF TRANSVERSE VIBRATIONS ALONG A STRETCHED STRING:

- The fundamental frequency of a string is inversely proportional to the length of the string, when the tension and linear density are constant.
  - $n \propto 1/l$  if T and m are kept constant or nl = constant.
- 2. The fundamental frequency of a string is directly proportional to the square root of the tension, when the length and linear density are constant.

 $n \propto \sqrt{T}$  if *l* and *m* are kept constant or  $n/\sqrt{T}$  = constant.

3. The fundamental frequency of a string is inversely proportional to the square root of its linear density, when the length and tension are constant.

 $n \propto 1/\sqrt{m}$  if *l* and *T* are constant or  $n\sqrt{m}$  is constant.

# VERIFICATION OF LAWS OF TRANSVERSE VIBRATIONS ALONG A STRETCHED STRING:

# SONOMETER

Description : The apparatus is shown in Fig. 1.7.

![](_page_44_Figure_17.jpeg)

It consists of a thin metallic wire stretched across two movable bridges A and B on the top of a hollow, wooden sounding box (S) about one metre long. One end of the wire is fixed to a peg at one end of the box. The other end of the wire passes over a smooth fixed pulley (P) and carries a weight hanger W. The length of the vibrating segment of the wire can be altered with the help of the movable bridges. The length of the vibrating segment can be measured by a scale fixed below the wire.

**I Law :** A suitable tension T is applied to the sonometer wire. A small paper rider is placed on the wire. A tuning fork of known frequency is excited and its stem is pressed on the sounding box. The length of the wire is adjusted until the paper rider placed at the centre of the vibrating segment is thrown off. The length of the vibrating segment (l) is measured. Now, the frequency of the wire is the same as the frequency of the tuning fork. The experiment is repeated with forks of different frequencies, keeping the tension the same. The results are tabulated as shown:

No.	Frequency of the fork (n)	Length of vibrating segment (l)	nl

The product *nl* will be constant, thus verifying the first law.

**II Law**:  $n \propto \sqrt{T}$  when *l* and *m* are kept constant. It is difficult to verify the second law directly. Choose a tuning fork of frequency *n*. The wire is stretched by a load of 2 kg. The length of the wire vibrating in unison with the fork is determined. The experiment is repeated for loads of 3, 4 and 5 kg for the same fork. The readings are tabulated as shown below :

Frequency of the fork = ..... (constant)

No.	Tension (T)	Length of vibrating segment l	$\frac{\sqrt{T}}{l}$

It is found that  $\sqrt{T}$  /l is a constant thereby verifying the law, indirectly.

As  $n \propto \sqrt{T} / l$  and as *n* is kept the same,  $\sqrt{T} / l$  is a constant.

**III Law :**  $n \propto 1/\sqrt{m}$  when *l* and *T* are constant. It is difficult to verify the third law directly. For a convenient tension (say 3 kg), the length of the wire that vibrates in unison with a given tuning fork is determined. The experiment is repeated for three wires of different materials for the *same tension* and for the *same fork*. The linear density (*m*) of each wire is determined by finding the mass of about 1 metre of the wire. The results are tabulated as shown below :

Tension = ...... (constant); Frequency of the fork = .........(constant)

No.	Linear density of the wire m	Length of the vibrating segment l	$l\sqrt{m}$

It is found that  $l\sqrt{m}$  is a constant, thereby verifying the third law indirectly.

As  $n \propto \frac{1}{l\sqrt{m}}$  and as *n* is kept the same,  $l\sqrt{m}$  is constant.

# **MEASUREMENT OF A.C. FREQUENCY USING SONOMETER FOR STEEL WIRE:**

# 1.10 A.C. Frequency Measurement using Sonometer

The frequency of the alternating current mains in the laboratory can be determined using a sonometer.

**Description :** A sonometer consists of a thin uniform wire stretched over two bridges on a wooden box (Fig. 1.11). One end of the wire is fixed to a peg. The other end of the wire passes over a pulley and carries a weight hanger. The

![](_page_46_Figure_4.jpeg)

length of the vibrating segment of the wire can be altered with the help of the movable bridges. The length of the vibrating segment can be measured by a scale fixed below the wire.

**Experiment :** A steel wire is mounted on a sonometer under suitable tension. An electromagnet is excited by the low voltage alternating current whose frequency is to be determined. The electromagnet is placed just above the sonometer wire. The wire is attracted twice in each cycle.

A small paper rider is placed on the wire. The length of the wire is adjusted until the paper rider placed at the centre of the vibrating segment is thrown off. The length of the vibrating segment (l) is measured. The experiment is repeated for different tensions. The readings are tabulated as shown below :

No.	Tension (T)	Length of the vibrating segment l	$\frac{\sqrt{T}}{l}$

The mean value of  $\frac{\sqrt{T}}{l}$  is found.

The mass per unit length of the wire is determined by finding the mass of a given length of the wire.

Calculation : The frequency of sonometer wire is

1

$$n = \frac{1}{2l} \sqrt{\frac{T}{m}}.$$

During both the positive peak and the negative peak of the A.C., the wire is pulled by the electromagnet. So the wire vibrates twice, for each cycle of the A.C.

The frequency of the A.C. supply is given by

$$f = \frac{n}{2}$$

Hence, the frequency of A.C. mains is calculated.

# FOR BRASS (OR) COPPER WIRE:

The ends of the secondary of a transformer are connected to the two ends A and B of the wire (Fig. 1.12). The P.D. at the ends of the transformer should be about 6 volts. The wire is set between the poles of a powerful horse shoe magnet or the opposite poles of two equal bar magnets, so that the magnetic field is in a horizontal plane and at right angles to the length of the wire.

![](_page_47_Picture_2.jpeg)

Fig. 1.12

The sonometer wire is subjected to a load of 150 gm. The wire is thrown into clear vibrations. The length of the wire is adjusted by moving the bridges C and D, till a single loop is seen between them with maximum

amplitude. The distance CD is measured with a metre scale. The experiment is repeated for loads of 200, 250, 300, 350, 400, 450 and 500 gm.

The length CD is measured in each case.

The frequency of the wire 
$$n = \frac{1}{2l} \sqrt{\frac{Mg}{m}}$$

In this arrangement, the frequency of the wire is equal to the frequency of the A.C. mains.

# **ULTRASONICS:**

The human ear is sensitive to sound waves in the frequency range from 20 to 20,000 Hz. This range is called audible range. Sound waves of frequency more than 20,000 Hz are called ultrasonics. These frequencies are beyond the audible limit.

# 1.12 Piezoelectric Effect

If one pair of opposite faces of a quartz crystal is subjected to pressure, the other pair of opposite faces develops equal and opposite electric charges on them (Fig. 1.13). The sign of the charges is reversed when the faces are subjected to tension instead of pressure. The electric charge developed is proportional to the amount of pressure or tension. This phenomenon is called *Piezoelectric effect*.

The effect is *reversible*, *i.e.*, if an electric field is applied across one pair of faces of the crystal, contraction or expansion occurs across the other pair.

![](_page_47_Figure_14.jpeg)

# **PRODUCTION OF ULTRASONIC WAVES BY PIEZOELECTRIC METHOD -PIEZOELECTRIC OSCILLATOR:**

**Principle :** This is based on the inverse piezoelectric effect. When a quartz crystal is subjected to an alternating potential difference along the electric axis, the crystal is set into elastic vibrations along its mechanical axis. If the frequency of the electric oscillations coincides with the natural frequency of the crystal, the vibrations will be of large amplitude. If the frequency of the electric field is in the ultrasonic frequency range, the crystal produces ultrasonic waves.

**Construction :** The circuit diagram is shown in Fig. 1.14. It is a base tuned oscillator circuit. A slice of quartz crystal is placed between the metal plates A and B so as to form a parallel plate capacitor with the crystal as the dielectric. This is coupled to the electronic oscillator through primary coil  $L_3$  of the transformer.

![](_page_48_Figure_3.jpeg)

Coils  $L_2$  and  $L_1$  of oscillator circuit are taken from the secondary of the transformer. The collector coil  $L_2$  is inductively coupled to base coil  $L_1$ . The coil  $L_1$  and variable capacitor  $C_1$  form the tank circuit of the oscillator.

**Working :** When the battery is switched on, the oscillator produces high frequency oscillations. An oscillatory e.m.f. is induced in the coil  $L_3$  due to transformer action. So the crystal is now under high frequency alternating voltage.

The capacitance of  $C_1$  is varied so that the frequency of oscillations produced is in resonance with the natural frequency of the crystal. Now the crystal vibrates with large amplitude due to resonance. Thus high power ultrasonic waves are produced.

# Advantages :

- 1. Ultrasonic frequencies as high as 500 MHz can be generated.
- 2. The output power is very high. It is not affected by temperature and humidity.
- 3. It is more efficient than magnetostriction oscillator.
- The breadth of the resonance curve is very small. So we can get a stable and constant frequency of ultrasonic waves.

# **Disadvantages :**

- 1. The cost of the quartz crystal is very high.
- Cutting and shaping the crystal is very complex.

# **PROPERTIES OF ULTRASONIC WAVES:**

- Ultrasonic waves vibrate at a frequency greater than the audible range for humans (20 hertz to 20 kilohertz).
- ✤ They have smaller wavelengths. As a result, their penetrating power is high.
- They cannot travel through vacuum.
- Ultrasonic waves travel at the speed of sound (330 m/s) in the medium. They have maximum velocity in a denser medium.
- ◆ In a homogeneous medium, they travel at a constant velocity.
- In low viscosity liquids, ultrasonic waves produce vibrations
- They undergo reflection, refraction and absorption.
- They have high energy content. They can be transmitted over a large distance without much loss of energy.
- They produce intense heat when they are passed through objects.
- Like sound waves, ultrasonic waves are longitudinal waves that produce alternate compressions and rarefactions.

# USES (OR) APPLICATIONS OF ULTRASONIC WAVES:

# Scientific Applications

1. Determination of the depth of sea : The ultrasonic waves are directed to the bottom of the sea. The time (t) that elapses between the emission of ultrasonic waves and the reception of the echo is recorded.

Depth of the sea = h = vt/2.

Here, v = the velocity of ultrasonic waves through sea-water.

2. Detection of aircraft, submarines etc : Because of their short wavelength, ultrasonic waves are highly directional. A piezoelectric quartz crystal oscillator is used for sending out a beam of ultrasonic waves. This is reflected back from an aircraft, if one comes in its way. The reflected beam is detected by a quartz receiver. The entire system is called SONAR from *sound navigation and ranging*.

**3.** Sound signalling : The highly directional ultrasonic sound beam is used for purposes of signalling to a distant ship.

# Industrial Uses :

1. Ultrasonic waves are used to detect cracks or flaws in metal structures.

**2. Ultrasonic soldering :** Ultrasonic solders are used for soldering aluminium coil capacitors, aluminium wires and plates without using any fluxes.

**3.** Ultrasonic welding : The properties of some metals change on heating. Therefore, they cannot be welded by electric or gas welding. In such cases, the metal sheets are welded together at room temperature using ultrasonic waves.

4. Ultrasonic drilling and cutting : Ultrasonics are used for making holes in very hard materials such as glass, diamond, gems and ceramics.

- 5. Ultrasonics are used for producing alloys of uniform composition.
- 6. Ultrasonics are used for cleansing clothes and parts of watches, etc.

# 7. Chemical Applications :

- (*i*) Ultrasonic waves are used to form stable emulsions of even immiscible liquids like *water and oil* or *water and mercury*. This finds an application in the preparation of photographic films, face creams, etc.
- (*ii*) They are used to liquefy gels like aluminium hydroxide in the same way as they are liquefied by shaking.
- (*iii*) They are used to coagulate fine solid or liquid particles in a gas ; for example, dust, smoke, mist, etc. Ultrasonics thus find use in collecting factory dust and purifying the air.
- *(iv)* Ultrasonics act like a catalytic agent and accelerate chemical reactions. Ultrasonic waves accelerate crystallisation.

# **Medical Applications :**

**1. Disease treatment :** Ultrasonic therapy can be used to treat diseases like bursites, abscesses, neuralgic and rheumatic pains, etc.

2. Surgical use : Kidney stones and brain tumours can be removed without shedding any blood using ultrasonic waves. Also, any tissue in our body can be cut selectively during an operation using ultrasonics.

**3. Diagnostic use :** Ultrasonics are used for detecting tumours and other defects in human body. State of breast cancer can be identified using ultrasonics in a non-destructive manner. Also, the twins or any defect in the growth of foetus can be identified using ultrasonics before delivery.

4. Extraction of broken teeth : Dentists use ultrasonic waves to properly extract broken teeth.

5. Sterilization : Ultrasonic waves can kill bacteria. Therefore, they are used for sterilisin milk.

6. Blood flow meters : Ultrasonic Doppler blood flow meters are used to study the bloc flow velocities in blood vessels of our body.

# **ACOUSTICS OF BUILDINGS:**

# 1.15 Reverberation

When a source produces sound waves inside a closed building, the waves are generally reflected repeatedly by walls, ceiling and other materials present in the room. The intensity of the sound wave decreases at every reflection and finally the sound becomes inaudible. So the listener receives (*i*) direct waves and (*ii*) reflected waves due to multiple reflections, at the various surfaces. There is also a time gap between the direct wave received by the listener and the waves received by successive reflection. Due to this, the sound persists for some time even after the source has stopped. This persistence of sound is called *reverberation*.

**Definition of Reverberation :** *Reverberation is the persistence of sound in an enclosure due to multiple reflections of sound at the walls after the source has ceased to emit sound.* 

**Definition of Reverberation time :** *The interval of time taken by sustained or continuous sound to fall in intensity to one millionth of its original value is called reverberation time (T).* 

It is the time taken for the sound to fall below the minimum audibility after the source stopped sounding.

The reverberation time will depend on the size of the room or the auditorium, the nature of the reflecting material on the wall and the ceiling and the area of the reflecting surfaces.

# Sabine's formula for Reverberation Time

Sabine's formula for reverberation time is

$$T = \frac{0.158V}{\Sigma \alpha A} \qquad \dots (1)$$

Here,

V = volume of the hall.  $\Sigma \alpha A = \alpha_1 A_1 + \alpha_2 A_2 + \alpha_3 A_3 + \alpha_4 A_4 + \alpha_5 A_5 +$ 

$$\alpha_1, \alpha_2, \alpha_3, \dots$$
 are the absorption coefficients of the surfaces of areas  $A_1, A_2, A_3, \dots$  respectively.

Eq. (1) shows that the reverberation time depends on volume and total absorbing power of the hall

# 1.16 Absorption Coefficient

**First Definition.** *The sound absorption coefficient* ( $\alpha$ ) *of a material is defined as the ratio of sound energy absorbed by it to the total energy incident on it.* 

Absorption coefficient ( $\alpha$ ) =  $\frac{\text{Sound energy absorbed by the surface}}{\text{Total sound energy incident on the surface}}$ 

Sabine chose 1 square metre of an open window as a standard unit of absorption since all sound waves falling on it pass through and can be said to be completely absorbed. This unit is called *sabine*. It is also called open window unit (O.W.U.).

**Second Definition.** *The absorption coefficient of a material is defined as the ratio of sound energy absorbed by it to that obsorbed by an equal area of an open window.* 

# 1.17 Acoustics of Buildings

The branch of Physics which deals with the design and construction of buildings with good acoustic properties is called "Acoustics of buildings." It deals basically with (*i*) reverberation control, (*ii*) noise insulation and reduction and (*iii*) sound distribution and absorption.

# **CONDITIONS FOR THE PERFECT ACOUSTICS:**

The following are the requirements of a good auditorium:

- 1. The sound should be sufficiently loud and intelligible in every part of the hall.
- Sound of each syllable should soon decay so that the succeeding syllable may be heard distinctly. There must be no confusion due to overlapping of syllables. This means that the auditorium must be free from excessive reverberation.
- 3. No echoes should be present.

- 4. There should not be *undesirable focussing* of sound in any part of the hall. There should not be any *zones of silence* or regions of poor audibility anywhere in the hall.
- 5. Interference, reflection and resonance effects should be avoided.
- 6. All extraneous noises must be shut out as far as possible. The boundaries should be sufficiently soundproof to exclude extraneous noise.
- 7. There should be no Echelon effect.
- 8. The quality of sound must be unaltered, *i.e.*, the relative intensity of the components of the complex sound must be preserved.

Fig. 1.15 shows the usual design of the hall.

![](_page_52_Figure_6.jpeg)

Fig. 1.15

It has a parabolic shape at the speaker's end *S*. As the speaker speaks, sound is reflected from the back almost as a parallel beam. So there is uniform distribution of sound intensity throughout the hall.

![](_page_53_Picture_0.jpeg)

# SCHOOL OF SCIENCE & HUMANITIES DEPARTMENT OF PHYSICS

**UNIT - V - Allied Physics I - SPH1217** 

# **GEOMETRICAL OPTICS AND PHYSICAL OPTICS**

# **INTRODUCTION:**

Optics is the branch of physics that studies the behaviour and properties of light, including its interactions with matter and the construction of instruments that use or detect it. In general, geometrical optics is based on the concept of rays. Whereas, physical optics deals with the explanation of physical phenomenon.

# LENS AND ITS TYPES:

A lens is an optical device which transmits and refracts light, converging or diverging the beam. Lenses are classified by the curvature of the two optical surfaces. A lens is biconvex (or double convex, or just convex) if both surfaces are convex; and a lens with two concave surfaces is biconcave (or just concave). If one of the surfaces is flat, the lens is plano-convex or plano-concave depending on the curvature of the other surface. A lens with one convex and one concave side is convex–concave or meniscus. It is this type of lens that is most commonly used in corrective lenses.

If the lens is biconvex or plano-convex, a collimated or parallel beam of light travelling parallel to the lens axis and passing through the lens will be converged (or focused) to a spot on the axis, at a certain distance behind the lens (known as the focal length). In this case, the lens is called a positive or converging lens. If the lens is biconcave or plano-concave, a collimated beam of light passing through the lens is diverged (spread); the lens is thus called a negative or diverging lens. The beam after passing through the lens appears to be emanating from a particular point on the axis in front of the lens; the distance from this point to the lens is also known as the focal length.

![](_page_54_Figure_6.jpeg)

Figure 4.3 Two types of lens.

# DEFECTS OF IMAGES (LENS) OR ABERRATION:

# One of the basic problems of lenses is the imperfect quality of images. The deviations in size, shape, position and colour in the actual images produced by a lens in comparison to the object are called aberrations produced by the lens.

According to geometric optics, the image of a point object formed in a lens is a point image. In reality, the image of a point object is not a point image, but it is spread in to a region in space both along and perpendicular to the axis of the lens. <u>The deviation of an optical image in size, shape and position</u> formed by a lens is known as aberration of an optical image.

The aberration of an image is not due to any defect in the construction of the lens, but it is due to the reasons mentioned below:

(1) The phenomenon of refraction in the lens and

(2) Variation of refractive index of the material of a lens with the wavelength of light.

# Monochromatic aberration:

The aberration of optical image even when monochromatic light is used is known as *monochromatic aberration*.

There are five different types of monochromatic aberrations. They are,

- (1) Spherical aberration
- (2) Coma
- (3) Astigmatism
- (4) Curvature of the field and
- (5) Distortion

# Chromatic aberration:

Aberration of optical images formed in a lens due to the variation of refractive index with the wavelength of light is known as *chromatic aberration*.

# SPHERICAL ABERRATION:

![](_page_55_Figure_17.jpeg)

# Figure (1) Spherical aberration in convex lens

The rays of light from the distant object after passing through the lens at the margin of the lens [known as marginal rays] converge at a point I<sub>m</sub> close to the lens. Similarly, the rays of light passing through a region close to the axis [known as paraxial rays] converge at a point I<sub>p</sub>, away from the lens. This results in an image that spreads into a region from I<sub>m</sub> to I<sub>p</sub> along the axis and from A to B perpendicular to the axis. This defect of the image due to the rays passing through different section of the lens, even with monochromatic light, is known as spherical aberration of the lens.

The spread of the image along the axis, [dx] is known as **longitudinal spherical aberration**. The image formed at AB is a circle with least diameter and at this position the best image is formed. This circle is called the circle of least confusion. The radius of the circle of least confusion measures lateral spherical aberration.

Note: The spherical aberration in a convex lens is taken to be positive as the marginal image is formed near the lens than the paraxial image. In the case of concave lens the spherical aberration is taken to be negative as the marginal image is formed to the right side of the paraxial image.

# **METHODS OF MINIMIZING SPHERICAL ABERRATION:**

- (1) By using stops: In this case, the stops used will either allow the paraxial rays or marginal rays. Usually the stop is used to avoid the marginal rays. This brings paraxial and marginal images close to one another thereby reducing the spherical aberration.
- (2) By the use of Plano-convex lens: In a lens, the deviation produced by the lens is minimum, when the deviation is shared equally between the two surfaces of the lens. This is achieved in a Plano-convex lens by arranging convex side facing the incident or emergent rays whichever are more parallel to the axis as shown in the following figure(2)

![](_page_56_Figure_6.jpeg)

(3) By the use of crossed lenses: It is theoretically known that the lenses have minimum spherical aberration when the parallel rays fall of the lens having their radii of curvature  $r_1$  and  $r_2$  bearing a ratio, which satisfies the following condition.

$$\frac{r_1}{r_2} = \frac{\mu(2\mu - 1) - 4}{\mu(2\mu + 1)} \dots \dots (1)$$

In the above equation,  $\mu$  is the refractive index of the material of the lens. For a lens of  $\mu = 1.5$ , the ratio

 $\frac{r_1}{r_2} = -\frac{1}{6}$ . A lens having its radii of curvature satisfying this condition is known as a **crossed lens**.

(4) By using two Plano-convex lenses separated by a suitable distance: When the two plano-convex lenses are separated at a suitable distance, the total deviation is divided equally between the two lenses and the total deviation is minimum. This reduces the spherical aberration to minimum. The necessary condition is derived as follows.

![](_page_56_Figure_12.jpeg)

With reference to figure (4), we can write,  $\angle BAK = \angle BF_2O_2 = \delta, \text{ Also, } \angle F_1BF_2 = \angle BF_2F_1 = \delta,$ So that  $F_1F_2 = F_1B = F_1O_2$  Or  $O_2F_1 = \frac{1}{2}O_2F_2.$  Since F<sub>2</sub> is the virtual object of the real image F<sub>1</sub> and using the lens formula for the second lens, we can write the equation,  $\frac{1}{v} - \frac{1}{u} = \frac{1}{f_2} - (2)$ 

In this equation  $u = f_1 - d \& v = \frac{f_1 - d}{2} - (3)$ 

Substituting for 'u' and 'v' and simplifying, we get,

$$\frac{2}{f_1 - d} - \frac{1}{f_1 - d} = \frac{1}{f_2} \Longrightarrow \frac{1}{f_1 - d} = \frac{1}{f_2} \Longrightarrow f_2 = f_1 - d$$
  
Or,  $d = f_1 - f_2 - \cdots - (4)$ 

Equation (4) gives the condition for minimum spherical aberration.

(5) By using suitable concave and convex lenses in contact: Since spherical aberration produced by convex lens is positive and that produced by a concave lens is negative, a suitable combination of convex and concave lens will minimize the spherical aberration.

# **CHROMATIC ABERRATION:**

Light waves of different colours have different velocities in a refracting medium. Every refracting medium, therefore has a different refractive index for each colour or wavelength. The refractive index is least for red colour and maximum for violet colour in the visible region. As the focal length of the lens depends upon the refractive index, it is different for different colours or wavelengths. As  $\mu_v > \mu_r$ , the focal length  $f_v < f_r$ . A single lens will, therefore, not form one image of an object point, but a series of coloured images at different distances from the lens, corresponding to each colour constituting the incident beam are formed. The variation of the image distance from the lens, with refractive index measures the **axial or longitudinal chromatic aberration**.

![](_page_57_Figure_8.jpeg)

Again the magnification of the image depends upon the focal length of the lens. The size of the image is different for different colours or wavelengths. The variation in the size of the image measures the **lateral chromatic aberration**. As shown in Fig. 21.1, the object *AB* gives the image  $A_1$   $B_1$  corresponding to violet light and  $A_2 B_2$  corresponding to the red light. The violet image is formed nearer to the lens than the red image. The distance  $B_1 B_2 = x$  measures the longitudinal chromatic aberration and the distance  $y = A_2 B_2 - A_1 B_1$  measures the lateral chromatic aberration. The images of intervening colours between violet and red lie between the images  $A_1 B_1$  and  $A_2 B_2$  and their size increases from violet to red.

It is clear that no one position the images are in sharp focus. The effect due to both the aberrations is known as **chromatic aberration**. The chromatic aberration of a lens for an object at infinity can be calculated as given below:

# ACHROMATISM OF THIN LENSES:

When two lenses are combined together in such a way that the image formed by the combination is free from chromatic aberration, then such a combination is called achromatic combination. The phenomenon of removal or minimization of chromatic aberration is known as achromatism. The achromatism is ideal if both the lateral and longitudinal chromatic aberrations are completely removed, i.e. the images of different colours are formed at the same position with same size. However, it is not possible to achieve the ideal achromatism. Usually, the combination is achromatized for size rather than for position. We shall now discuss two cases for achromatic combination of two lenses.

- (1) Achromatism for two thin lenses in contact
- (2) Achromatism for two thin lenses separated by a distance

# ACHROMATIC COMBINATION OF TWO THIN LENSES IN CONTACT:

For achromatic doublet, consider two thin lenses of different meterials placed in contact with each other. For the combination of two lenses to be achromatic, we will find the condition for this lens combination to have the focal length independent of the colours. The focal length f of a thin lens is given by

$$\frac{1}{1}$$
 (1, 1)

 $\left(\frac{1}{R_1} - \frac{1}{R_2}\right) = \frac{1}{(\mu - 1)f}$ 

or

.....(2)

where  $\mu$  is the refractive index of the lens,  $R_1$  and  $R_2$  are the radii of curvature of the two surfaces of the lens. If  $\delta f$  is the change in focal length f corresponding to a change  $\delta \mu$  in the refractive index  $\mu$ , then differentiating eqn. (1) with respect to f, we get

$$-\frac{\delta f}{f^2} = \delta \mu \left(\frac{1}{R_1} - \frac{1}{R_2}\right)$$
  
From eqn. (2), we have  $\frac{-\delta f}{f^2} = \frac{\delta \mu}{(\mu - 1)} \cdot \frac{1}{f}$  .....(3)

But  $\frac{\delta\mu}{(\mu-1)} = \omega$ , the dispersive power of the lens between the two colours for which the difference in refractive index is  $d\mu$  and mean refractive index is  $\mu$ .

Negative sign in the left hand side indicates that the focal length of the lens decreases as refractive index increases. If  $f_1$  and  $f_2$  be the mean focal lengths of two thin lenses of combination, and  $\omega_1$  and  $\omega_2$  be the dispersive powers between the two colours for which the combination is to be achromatised. As the lenses are in contact, the focal length of the combination is given by

Differentiating eqn. (5) partially, we get  $-\frac{\delta F}{F^2} = -\frac{\delta f_1}{f_1^2} - \frac{\delta f_2}{f_2^2}$ 

But from eqn. (4), we have 
$$-\frac{\delta f_1}{f_1^2} = \frac{\omega_1}{f_1}$$
 and  $-\frac{\delta f_2}{f_2^2} = \frac{\omega_2}{f_2}$   $\therefore$   $-\frac{\delta F}{F^2} = \frac{\omega_1}{f_1} + \frac{\omega_2}{f_2}$  .....(6)

For achromatic combination,  $\delta F = 0$ , that is, F should not change with colour.

This is the required condition for achromatic doublet. According to this condition, (i) The ratio of the focal lengths of the two lenses is numerically equal to the ratio of the dispersive powers of their materials and (ii) Since  $\omega_1$  and  $\omega_2$  are both positive, the focal lengths of the two lenses must be of opposite sign, that is, if one lens is convex, the other should be concave as shown in fig. 6.

![](_page_59_Figure_9.jpeg)

# ACHROMATISM FOR TWO THIN LENSES SEPARATED BY A DISTANCE:

It is possible to minimise chromatic aberration by using two convex lenses of the same material provided they are separated by a finite distance. To find the condition of minimum chromatic aberration, consider two convex lenses of focal lengths  $f_1$  and  $f_2$  are placed coaxially at a distance d apart. The equivalent focal length F of the combination is given by

$$\mathbf{d} = \frac{\mathbf{f}_1 + \mathbf{f}_2}{2}$$

# Thus, for minimum chromatic aberration, the distance between the two lenses of the same material must be equal to half the sum of their focal lengths.

# **PHYSICAL OPTICS:**

Physical optics deals with the explanation of physical phenomenon such as Interference, Diffraction and Polarisation.

# **INTERFERENCE:**

When two light waves of same frequency and of nearly the same amplitude travel through the same region of space at the same time, then due to their superposition, the resultant intensity at different points of the medium undergoes change from point to point. The change in the intensity of light in a medium from point to point is called interference of light.

Thus, the modification or change in the uniform distribution of light intensity in a medium due to superposition of two light waves is called interference of light.

# AIR-WEDGE OR AIR-WEDGE SHEARING INTERFEROMETER:

- The air-wedge shearing interferometer is probably the simplest type of interferometer designed to visualize the disturbance of the wave front after propagation through a test object.
- This interferometer is based on utilizing a thin wedged air-gap between two optical glass surfaces.
- This interferometer consists of two optical glass wedges pushed together and then slightly separated from one side to create a thin air-gap wedge. This air-gap wedge has a unique property like very thin (micrometer scale) and perfect flatness.

![](_page_60_Figure_11.jpeg)

# DETERMINATION OF DIAMETER (OR) THICKNESS OF A THIN WIRE BY AIR WEDGE DIFFRACTION:

![](_page_61_Figure_1.jpeg)

S	-	Source (Sodium vapour light)
L	-	Condensing lens (convex lens)
G	-	Glass plate inclined at 45°
L1, L2 -		Transparent plane glass plates

Fig. 1 Experimental arrangement

![](_page_61_Figure_4.jpeg)

Fig. 2 Fringe pattern

### Aim

To determine the thickness of a thin wire using air wedge method

### Apparatus required

Two optically plane rectangular glass plates, thin wire, travelling microscope, reading lens, sodium vapour lamp, condensing lens with stand, wooden box with glass plate inclined at 45°.

#### Formula

# 1. The thickness of the wire, $t = \frac{l\lambda}{2R}m$

# 2. Wedge angle, $\theta = \frac{t}{1}$ radians

Where  $\boldsymbol{\lambda}$  is the wavelength of sodium light in m

- $\ell\,$  is the distance of the wire from the edge of contact (I) in m
- $\boldsymbol{\beta}$  is the width of one fringe in m

#### Procedure

The experimental arrangement and ray diagram are shown in Fig. 1. Two optically plane glass plates are placed one over other and tied together by means of a rubber band at one end. A thin wire is inserted between the plates at the other end. Now a wedge shaped air film is formed between the two glass plates. The slide system is kept on the platform of a travelling microscope. The light from a sodium vapour lamp is rendered parallel with a condensing lens and is made to incident on a plane glass sheet held over the wedge at an angle of 45° with the vertical. The light falling on the sheet is partially reflected which is in turn incident normally on the air wedge. Adjusting the arrangement properly, the microscope field of view is made bright to the maximum extent. The microscope is moved vertically up and down till parallel fringes are visible which are located on the surface of the air film (Fig.2). By moving the microscope in a horizontal direction, the cross-wires of the microscope are set on one of the dark (n<sup>th</sup>) fringes in the pattern. Its position is noted down in the horizontal scale.

The microscope is moved further using the tangential screw along the length of the air film counting the number of dark fringes. After counting 2 dark fringes, the cross wire is coincided with the  $(n+2)^{nd}$  fringe and its position is noted. The measurements are repeated similarly for every alternate dark fringe and are noted. The width of 10 dark fringes is calculated from the table and the mean width of 10 fringes is averaged out. From this, the fringe width  $\beta$  is calculated. The length of the air film is measured as the distance between the line of contact and the inner edge of the wire. The measurement can be done using the travelling microscope or with the calibrated scale.

#### To find the fringe width (β)

#### Least count = 0.001cm

Order	Mic	roscope	e reading	9	Order	Mic	croscope	reading	3	Width
of	M.S.R	V.S.C	V.S.R	T.R	of	M.S.R	V.S.C	V.S.R	T.R	of 25
fringes					fringes					fringes
	cm	div	cm	cm		cm	div	cm	cm	cm
n					n+25					
n+5					n+30					
n+10					n+35					
n+15					n+40					
n+20					n+45					

#### Mean width of 25 fringes =

#### Calculations

Wavelength of sodium light (λ)	=	m
Distance of the wire from the edge of contact (I)	=	m
Mean width of 25 fringes	=	m
Width of one fringe (β)	=	m

The thickness of the wire, t =  $\frac{\ell\lambda}{2\beta}$ 

# THEORY OF TRANSMISSION GRATING - NORMAL INCIDENCE:

# Diffraction grating

An arrangement consisting of a large number of equidistant parallel narrow slits of equal width separated by equal opaque portions is known as a diffraction grating.

The plane transmission grating is a plane sheet of transparent material on which opaque rulings are made with a fine diamond pointer. The modern commercial form of grating contains about 6000 lines per centimetre.

The rulings act as obstacles having a definite width 'b' and the transparent space between the rulings act as slit of width 'a'. The combined width of a ruling and a slit is called grating element (e). Points on successive slits separated by a distance equal to the grating element are called corresponding points.

### Theory

MN represents the section of a plane transmission grating. AB, CD, EF ... are the successive slits of equal width *a* and BC, DE ... be the rulings of equal width b (Fig. 5.21). Let e = a + b.

Let a plane wave front of monochromatic light of wave length  $\lambda$  be incident normally on the grating. According to Huygen's principle, the points in the slit AB, CD ... etc act as a source of secondary wavelets which spread in all directions on the other side of the grating.

![](_page_63_Figure_8.jpeg)

Fig 5.21 Diffraction grating

Let us consider the secondary diffracted wavelets, which makes an angle  $\theta$  with the normal to the grating.

The path difference between the wavelets from one pair of corresponding points A and C is  $CG = (a + b) \sin \theta$ . It will be seen that the path difference between waves from any pair of corresponding points is also  $(a + b) \sin \theta$ 

The point P1 will be bright, when

 $(a + b) \sin \theta = m \lambda$  where m = 0, 1, 2, 3

In the undiffracted position  $\theta = 0$  and hence  $\sin \theta = 0$ .

 $(a + b) \sin \theta = 0$ , satisfies the condition for brightness for m = 0. Hence the wavelets proceeding in the direction of the incident rays will produce maximum intensity at the centre O of the screen. This is called zero order maximum or central maximum.

If  $(a + b) \sin \theta_1 = \lambda$ , the diffracted wavelets inclined at an angle  $\theta_1$  to the incident direction, reinforce and the first order maximum is obtained.

Similarly, for second order maximum,  $(a + b) \sin \theta_2 = 2\lambda$ 

On either side of central maxima different orders of secondary maxima are formed at the point P1, P2.

In general,  $(a + b) \sin \theta = m \lambda$  is the condition for maximum intensity, where m is an integer, the order of the maximum intensity.

$$\sin \theta = \frac{m\lambda}{a+b}$$
 or  $\sin \theta = \text{Nm}\lambda$ 

When white light is used, the diffraction pattern consists of a white central maximum and on both sides continuous coloured images are formed.

where N = 1/a+b, gives the number of grating element or number of lines per unit width of the grating.

# DETERMINATION OF WAVELENGTH OF MONOCHROMATIC SOURCE USING A GRATING BY NORMAL INCIDENCE:

![](_page_64_Figure_12.jpeg)

The wavelength of a spectral line can be very accurately determined with the help of a diffraction grating and spectrometer.

Initially all the preliminary adjustments of the spectrometer are made. The slit of collimator is illuminated by a monochromatic light, whose wavelength is to be determined. The telescope is brought in line with collimator to view the direct image. The given plane transmission grating is then mounted on the prism table with its plane is perpendicular to the incident beam of light coming from the collimator. The telescope is slowly turned to one side until the first order diffraction image coincides with the vertical cross wire of the eye piece. The reading of the position of the telescope is noted (Fig. 5.22).

Similarly the first order diffraction image on the other side, is made to coincide with the vertical cross wire and corresponding reading is noted. The difference between two positions gives 2 $\theta$ . Half of its value gives  $\theta$ , the diffraction angle for first order maximum. The wavelength of light is calculated from the equation  $\lambda = \sin \theta / Nm$ . Here N is the number of rulings per metre in the grating.

# DETERMINATION OF WAVELENGTH OF MERCURY LINES (OR) WHITE LIGHT USING A GRATING BY NORMAL INCIDENCE:

Monochromatic light is now replaced by the given source of white light. The source emits radiations of different wavelengths, then the beam gets dispersed by grating and a spectrum of constituent wavelengths is obtained as shown in Fig 5.23.

![](_page_65_Figure_5.jpeg)

Fig 5.23 Diffraction of white light

knowing N, wave length of any line can be calculated from the relation

$$\lambda = \frac{\sin\theta}{Nm}$$

# Difference between interference and diffraction

	Interference	Diffraction
1.	It is due to the superposition of secondary wavelets from two different wavefronts produced by two coherent sources.	It is due to the superposition of secondary wavelets emitted from various points of the same wave front.
2.	Fringes are equally spaced.	Fringes are unequally spaced.
3.	Bright fringes are of same intensity	Intensity falls rapidly
4.	Comparing with diffraction, it has large number of fringes	It has less number of fringes.

# **POLARISATION:**

A light wave that is vibrating in more than one plane is referred to as unpolarized light. Polarized light waves are light waves in which the vibrations occur in a single plane. The process of transforming unpolarized light into polarized light is known as polarisation.

![](_page_66_Figure_4.jpeg)

# **OPTICAL ACTIVITY:**

When a plane polarised light is made to pass through certain substances, the plane of polarisation of the emergent light is not the same as that of incident light, but it has been rotated through some angle. This phenomenon is known as optical activity. The substances which rotate the plane of polarisation are said to be optically active. Examples : quartz, sugar crystals, turpentine oil, sodium chloride etc.

Optically active substances are of two types, (i) Dextro-rotatory (right handed) which rotate the plane of polarisation in the clock wise direction on looking towards the source. (ii) Laevo - rotatory (left handed) which rotate the plane of polarisation in the anti clockwise direction on looking towards the source.

Light from a monochromatic source S, is made to pass through a polariser P. The plane polarised light is then made to fall on an analyser A, which is in crossed position with P. No light comes out of A. When a quartz plate is inserted between the polariser and analyser some light emerges out of the analyzer A (Fig. 5.32). The emerging light is cut off again, when the analyzer is rotated through a certain angle.

This implies that light emerging from quartz is still plane polarised, but its plane of polarisation has been rotated through certain angle.

![](_page_67_Figure_0.jpeg)

The amount of optical rotation depends on :

- thickness of crystal
- density of the crystal or concentration in the case of solutions.
- wavelength of light used
- the temperature of the solutions.

# SPECIFIC OPTICAL ROTATORY POWER (OR) SPECIFIC ROTATION:

The capacity of a substance to cause rotation of the plane of polarization when plane-polarized light is passed through it.

The term specific rotation is used to compare the rotational effect of all optically active substances.

Specific rotation for a given wavelength of light at a given temperature is defined as the rotation produced by one-decimeter length of the liquid column containing 1 gram of the active material in 1cc of the solution.

If  $\theta$  is the angle of rotation produced by *l* decimeter length of a solution of concentration C in gram per cc, then the specific rotation S at a given wavelength  $\lambda$  for a given temperature *t* is given by

$$S = \frac{\theta}{l.c}$$

The instrument used to determine the optical rotation produced by a substance is called polarimeter.

Sugar is the most common optically active substance and this optical activity is used for the estimation of its strength in a solution by measuring the rotation of plane of polarisation.

# **POLARIMETER:**

- A polarimeter is a scientific instrument used to measure the angle of rotation caused by passing polarized light through an optically active substance.
- Some chemical substances are optically active and polarized (uni-directional) light will rotate either to the left (counter-clockwise) or right (clockwise) when passed through these substances.
- The amount by which the light is rotated is known as the angle of rotation. The angle of rotation is basically known as observed angle.

# DETERMINATION OF SPECIFIC ROTATORY POWER OF A SOLUTION USING THE POLARIMETER:

Polarimetry is an instrumental analytical method using rotation of polarized light by some substances as a measure of their concentration in a solution. The instrument used is called a polarimeter. When it is adapted for measuring quality of sugar the name saccharimeter is used. In both instruments it is the rotation of polarized light by a substance in a solution which is measured. Usually, it is only one instrument which has two interchangeable scales, one labelled in angular degrees <sup>0</sup>, the other in units <sup>0</sup>Z, named International Sugar Scale (I.S.S.).

### Polarimeter

Figure 1 shows a principle of a polarimeter set up and its main components together with their function. Unpolarized light from the light source is first polarized. This polarized light passes through a sample cell. If an optical active substance is in a sample tube, the plane of the polarized light waves is rotated. The rotation is noticed by looking through the analyser as a change in intensity of illumination. To reach the same illumination as was without an optical active sample the analyser must be turned around for an angle. Readings are taken in degrees (angle)  $\alpha$  or sugar degrees <sup>0</sup>Z.

![](_page_68_Figure_8.jpeg)

Figure 1: a) principle of a polarimeter set up

b) components of a polarimeter

Figure 3 shows the passage of polarized light through a sample tube, which is positioned between polarizer and analyzer. The length of the tube is one of important parameters to be fixed if the measurements were comparable.

![](_page_69_Picture_1.jpeg)

Figure 3: Rotation of polarized light plane by a sample with an optical active substance and the influence of the length of sample tube

### **Optical Rotation**

Certain compounds, mostly organic (notably those containing asymmetric carbon atoms) rotate the plane of polarized light. The phenomenon is called optical rotation and such substances optically active compounds.

Measuring angle of rotation the concentration of a substance in a solution is determined.

The measured angle of rotation depends upon many variables:

- The type or nature of sample (example: sugar solution)
- Concentration of the optical active components
- The length of the sample tube
- The wavelength of the light source
- Temperature of the sample

We describe the nature of a sample by introducing the specific optical rotatory power (or specific rotation) of a substance, defined as

$$[\alpha]^{\Theta}_{\lambda} = \frac{\alpha}{\gamma I}$$

in SI units: rad m<sup>2</sup> kg<sup>-1</sup>

(Notice:  $2\pi$  rad = 360 ° (deg of angle))

where  $\alpha$  is the angle of rotation in rad,  $\gamma$  is the mass concentration in kg/m<sup>3</sup>, and *I* is the length of the sample tube in m. Specific rotation is determined at a specified temperature  $\Theta$  (usually 20 °C) and a wavelength of light source (usually sodium lamp with its D line at 589 nm).

Some substances rotate the light to the right (or clockwise) as viewed looking towards the light source, we sign this rotation and  $\alpha$  as +, some to the left (or anticlockwise), signing  $\alpha$  as -.

In practical measurements readings are taken at different units:  $\alpha$  in <sup>0</sup> (deg),  $\gamma$  in g/cm<sup>3</sup>, / in dm and so  $[\alpha]_{D}^{20^{\circ}}$  is usually tabulated in <sup>0</sup> cm<sup>3</sup>/g dm.

Substance	Specific rotation		
in a solution H <sub>2</sub> O solvent	$[\alpha]_D^{20^0}$		
	[ <sup>0</sup> cm <sup>3</sup> /g dm ]		
sucrose	+ 66.54		
glucose	+ 52.74		
fructose	- 93.78		
maltose	+ 137.5		
lactose	+ 55.3		
dextrose	+ 194.8		