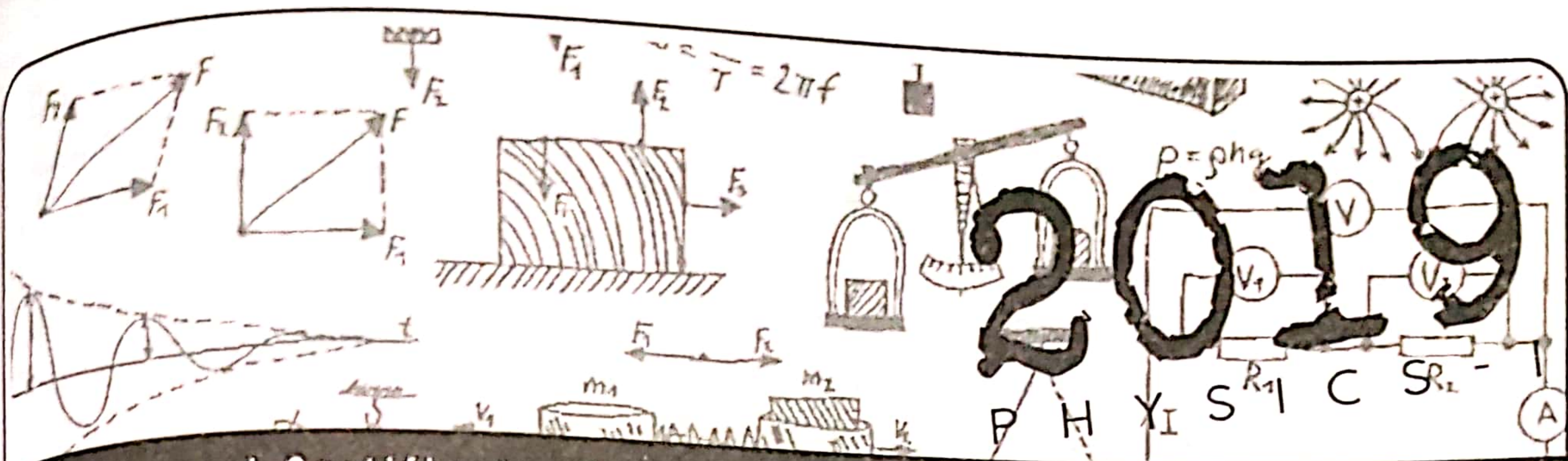




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General Certificate of Education (Adv. Level) Examination

01. Which of the following is not a fundamental unit?

- (1) m (2) J (3) cd (4) K (5) mol

Unit and Dimesions 01

Out of the following, which one does not represent a basic unit of SI? (2012)

- (1) m 2) N 3) kg 4) s 5) K

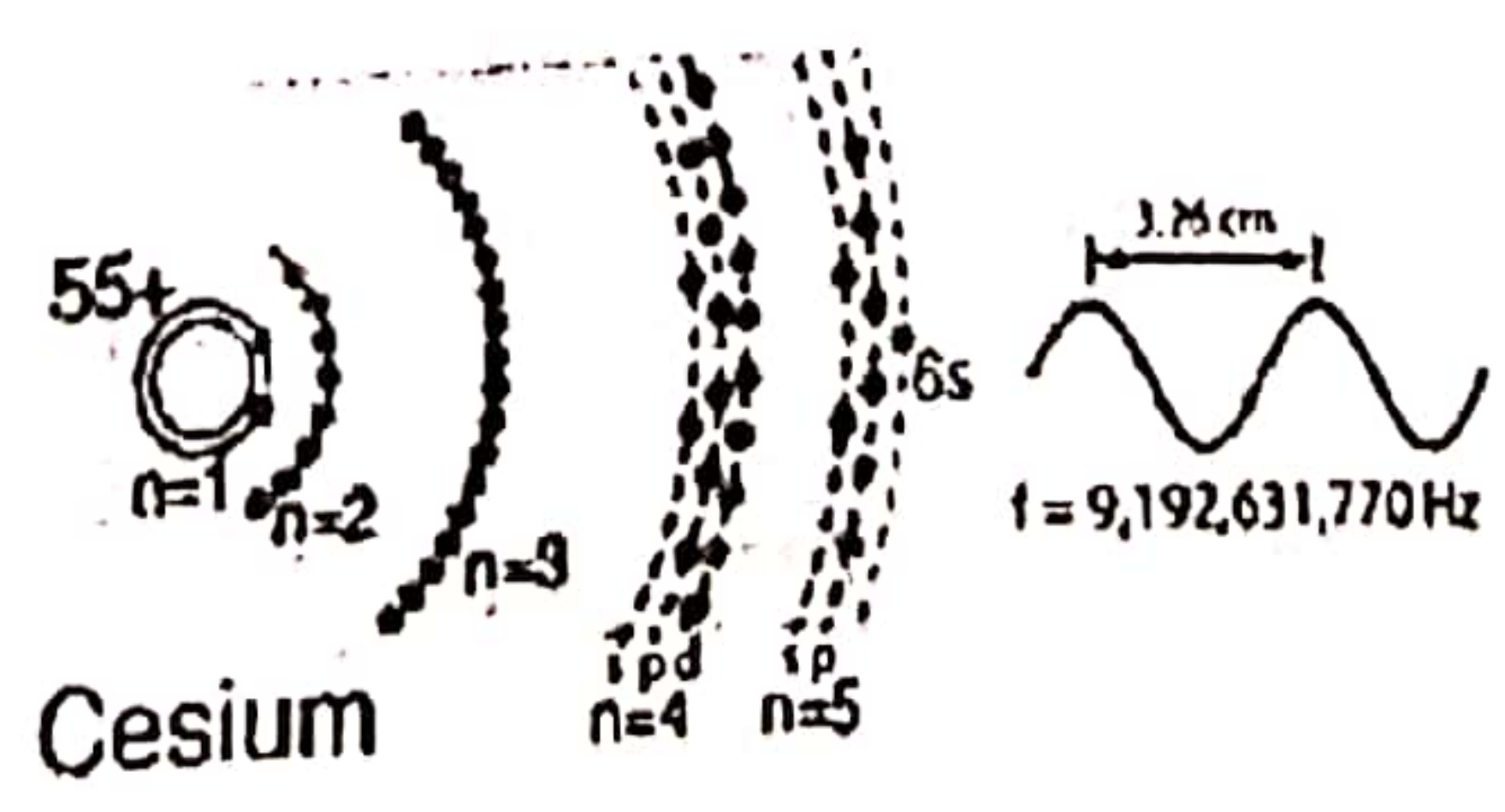
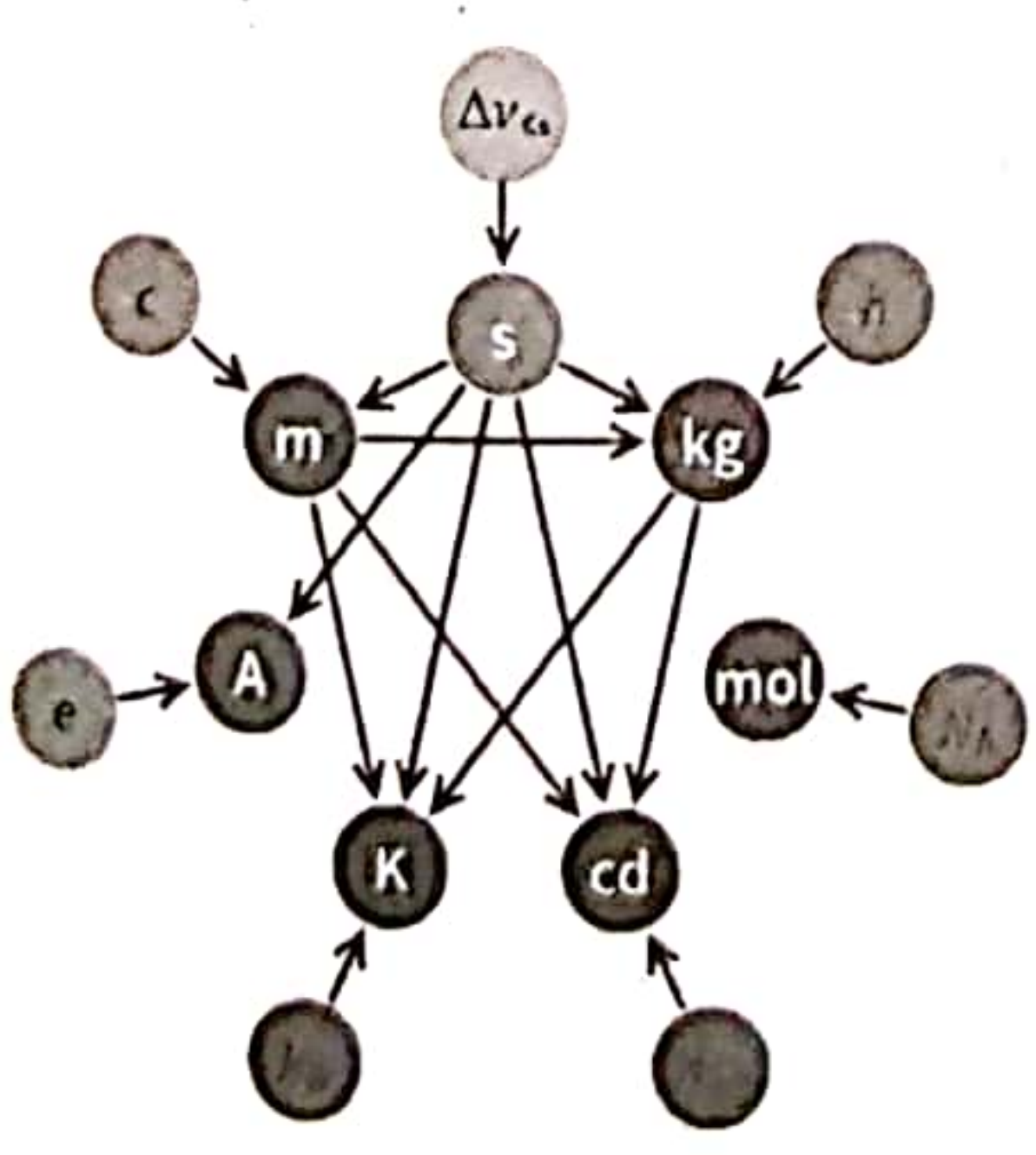
You very well know the basic quantities and their units of international unit system.

Basic Quantity	SI unit	Symbol
Length	meter	m
Mass	kilogram	kg
Time	second	s
Electric Current	Ampere	A
Temperature	Kelvin	K
Amount of matter	mole	mol
Light Intensity	candela	cd

Other than candela, the other units are familiar to us. The shortened form SI was made from the French word of 'Système International d'unités' which means International System of Units in English.

On 20th May 2019, the international committee of weights and measurements has interpreted the above units again. According to the newest interpretations, even though the values of the units are not changed, all the basic units are being interpreted only using the numerical values of the basic constants of the universe such as the speed of light (c), the charge of an electron (e), Planck's constant (h), Boltzmann's constant (k), Avogadro constant (N_A) and the transferring frequency of two certain energy levels of ¹³³ Cs (Cesium) atom.

Second (s)- The time taken for the radiation amount of 9.192631770 X 10⁹ emission in the transfer of shown two energy levels of ¹³³ Cs atom is interpreted as 1 s. These radiations belong to the microwave region. Why do we use ¹³³ Cs for this purpose?



^{133}Cs is being used in atomic clocks most of the time. The only stable isotope of Cesium atoms is ^{133}Cs . If we have taken another atom, then there can be difficulties in measuring the frequency correctly and precisely of the emitted radiation due to the transfers from other isotopes. The transfer that is happening in ^{133}Cs is shown in the above figure.

How do we measure time? Actually, what is time? From the ancient times, people have measured time using different methods. The time was interpreted by using many methods such as the distance of the shadow created from a specific object by the sun, the oscillations of different pendulums, the revolution by the earth around its axis etc.

After some time, scientists got to know that time can be very precisely calculated by the transfers between two energy levels of an atom. They chose ^{133}Cs atom as the best atom. The main reason for the selection was the time taken to do 9.192631770×10^9 transfers was very similar to the familiar solar second value. But the time interpretation based on the earth rotation of its own axis or around the sun is not correct. Because their travelling locus are not very certain. But by using atomic clocks like ^{133}Cs , the time can be measured correctly and precisely.

9.192631770×10^9 is the frequency of the emitted microwave radiation. That means its unit is Hz automatically (the amount of radiation emitter per second). From this, the second can be just interpreted.

Meter (m)- The newest interpretation is the oldest one. That means a meter is the distance that the light travels during a period of $1/2.99792458 \times 10^8$ from a second in a vacuum. 2.99792458×10^8 is the speed of light at a vacuum. The speed of light can be very accurately measured by the scientists.

Kilogram (kg)- The previous interpretation is the international prototype of the kilogram which is the mass of the cylinder made from Platinum and Iridium. The new interpretation is based on the measured value of the Plank's constant which is $6.62607015 \times 10^{-34}$. The unit of h is Js. That means $\text{kgms}^{-2}\text{ms} = \text{kgm}^2\text{s}^{-1}$. Therefore, by using the above interpretation of m and s for the usage of the h value, kg can be interpreted. If a mass at rest has a similar photon of energy with a frequency $1.356392489652 \times 10^{50}$ Hz, then that means the object has a mass equal to 1 kg.

$$m_0 c^2 = hf,$$

$$m_0 = \frac{hf}{c^2} = (6.62607015 \times 10^{-34} \times 1.356392489652 \times 10^{50}) / (2.99792458 \times 10^8)^2$$

$$m_0 = 1 \text{ kg}$$

Ampere (A) – The previous interpretation; When two straight parallel wires of an infinite length with negligible cross-sectional area are kept at 1m away at a distance in a vacuum and there are equal current flows in them, if the force per unit length of a conductor is $2 \times 10^{-7} \text{ Nm}^{-1}$, then the current which flows is equal to one Ampere. The new interpretation is very simple. There is no need for the values of force, distance and μ_0 . According to the interpretation of 2019, Ampere is interpreted from the basic charge of electric charge. The charge of an electron is $1.602176634 \times 10^{-19} \text{ C}$ ($\text{C} = \text{As}$). That means if $1.602176634 \times 10^{-19}$ charges are flown in $1.602176634 \times 10^{-19}$ time, then the current is one Ampere.

Kelvin (K)- The interpretation of Kelvin was subjected to a fundamental change. Previous interpretation; One Kelvin is $1/273.16$ times of the thermometric temperature of the water's triple point.

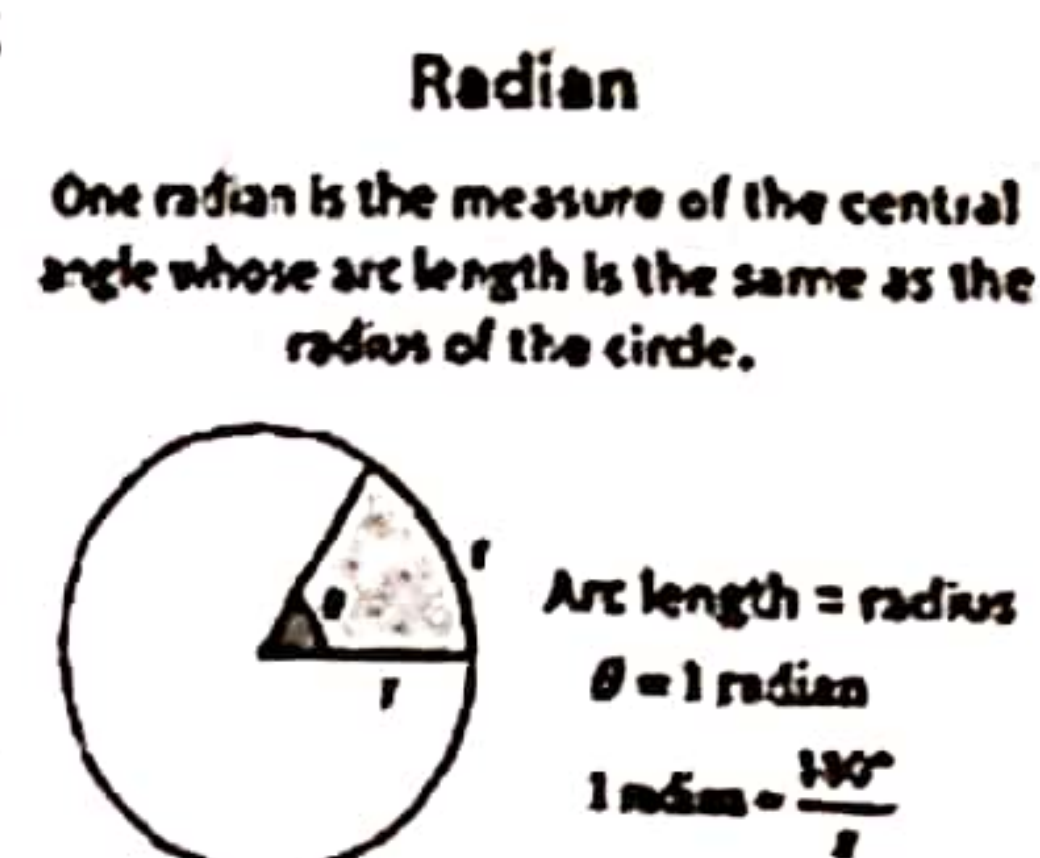
The new interpretation; The value of Boltzmann constant k is $1.380649 \times 10^{-23} \text{ JK}^{-1} = \text{kgm}^2\text{s}^{-2}\text{K}^{-1}$. The units of kg, m and s have been interpreted above. Therefore, using them we can interpret Kelvin. By equating hf value to kT , T can be interpreted. Both hf and kT are relations that gives energy.

Mole (mol)- The mole is interpreted from Avogadro's number which is $6.02214076 \times 10^{23}$. In a mole there

are $6.02214076 \times 10^{23}$ of elementary entities. These elementary entities can be atoms, molecules, ions, electrons or any other particle.

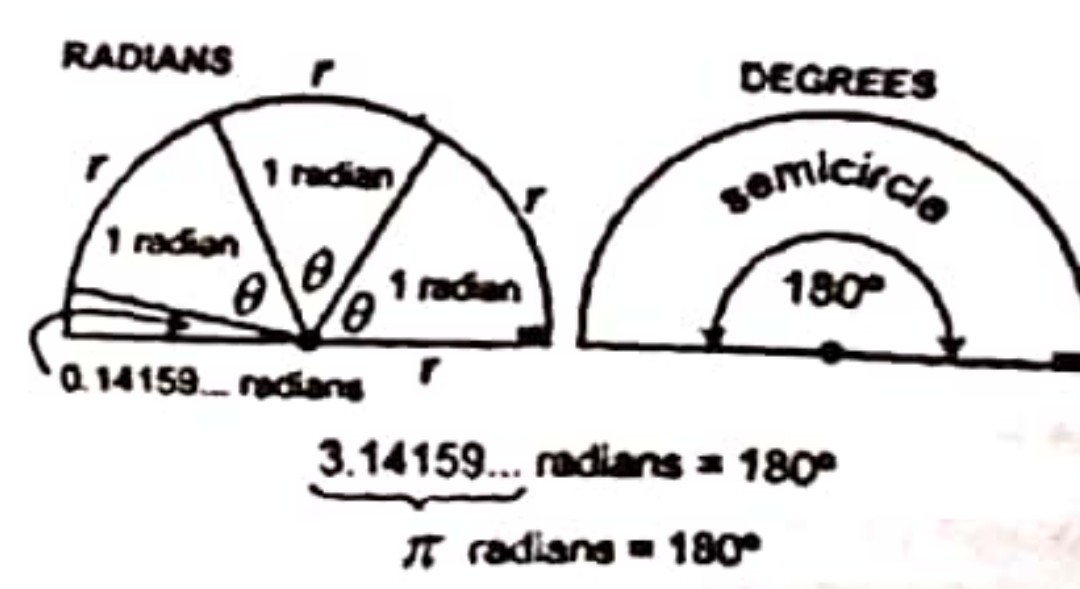
Candela (cd) – The interpretation of this has not been changed. If the radiation intensity of a source which emits monochromatic light with frequency 540×10^{12} Hz to a certain direction is $1/683$ W per steradian, then the luminous intensity of that source is one candela. The word candela may have been based from the candle. When there are no electric bulbs, we used to use candles. The frequency of 540×10^{12} Hz has been selected because the light (yellowish green) associated with that frequency (or the related wavelength 555 nm) gives a maximum reaction to our eyes. $1/683$ W has been taken due to this reason. If the emitted power for a solid angle with a steradian is $1/683 \text{ Js}^{-1}$, then the total power emitted from the surface is $1/683 \times 4\pi$. This is about 18.4 mW.

According to the interpretation, the luminous intensity from a standard candle is taking this value. Back in 1860, the first standard candle and candle power have been interpreted. One candle power has been interpreted as the total light emitting from a candle that is burning with 7.8 g per hour where the candle has been made from spermaceti wax (taken from the head of a whale) with a mass of $1/6$ pounds (76 g). The total light power from such a candle is nearly 18.4 mW. $1/683$ value has been obtained using this.

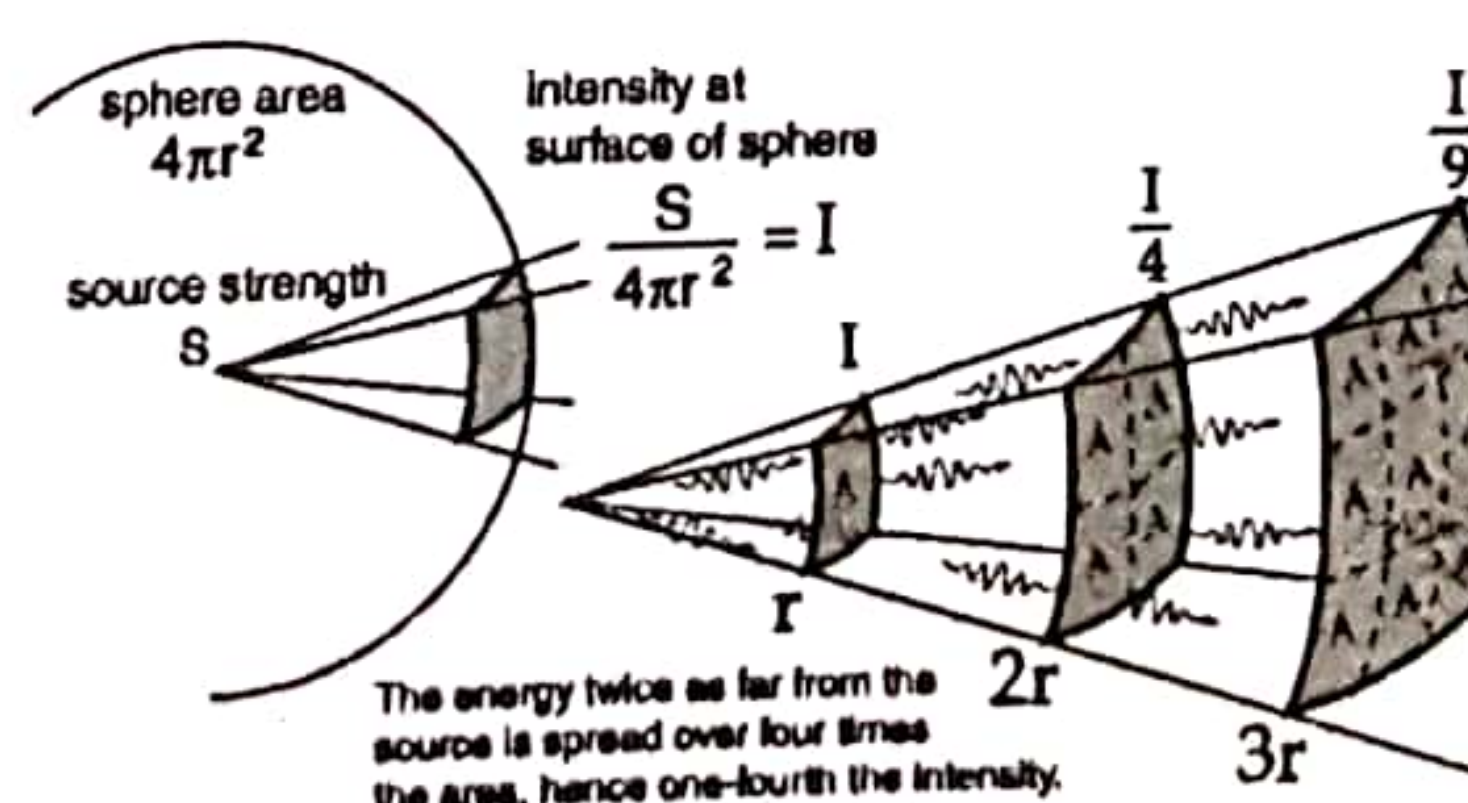


There can be a certain doubt of taking cd as a basic unit. To measure radiation intensity, we can however use Wm^{-2} or Wsterad^{-1} . Why then candela has been interpreted? Candela measures the luminous intensity. That means the sensation we feel in our eyes. The visual sensation is not solely dependent upon the radiation intensity. The sensation is dependent upon the frequency of the light. That is why a certain frequency has been selected which gives the maximum sensation to the eye.

You know that from radians are being used to measure plane angles (2D). The quantities of in rotation motion that are angular displacement, angular velocity, angular acceleration are measured in rad, rads^{-1} and rads^{-2} respectively.



As shown in the figure one radian can be interpreted like this way. If the arc is being selected as r in a circle with a radius r , then the angle subtended on the centre is one radian. The reason for calling this unit as a radian may be because all the lengths are equal to the radius. This is what I feel. If the angle from an arc length of r is one radian, then the radian amount from the total circumference length of $2\pi r$ should be $2\pi (1/r \times 2\pi r)$. Therefore, 2π radian should be equal to 360° . By this, if the angle is θ from any arc length of S , then you can write as $S = r\theta$. The solid angle is an angle that is made in three dimensions. It is measured in steradians (square radians). In Greek, the meaning of 'steros' is equal to solid in meaning. Steradian can be interpreted like the radian. The subtended solid angle on the centre by an area of r^2 of the surface of a sphere of radius r is one steradian. Look at the figure. The total surface area of a sphere is $4\pi r^2$.



Therefore, the solid angle subtended from an area of $4\pi r^2 = 1/r^2 \times 4\pi r^2 = 4\pi$

If the solid angle Ω is subtended by an area of A , then $\Omega = 1/r^2 \times A = A/r^2$. Likewise, $S = r\theta$, A can be written as $A = r^2\Omega$

J and N are not basic units. They can be expressed using basic units.

2. The dimensions of the gravitational constant G are given by
 (1) $L^2M^{-1}T^{-1}$ (2) L^2M^{-2} (3) $L^2M^{-2}T^{-1}$ (4) $L^3M^{-1}T^{-2}$ (5) $L^3M^{-2}T^{-2}$

01

Unit and Dimensions

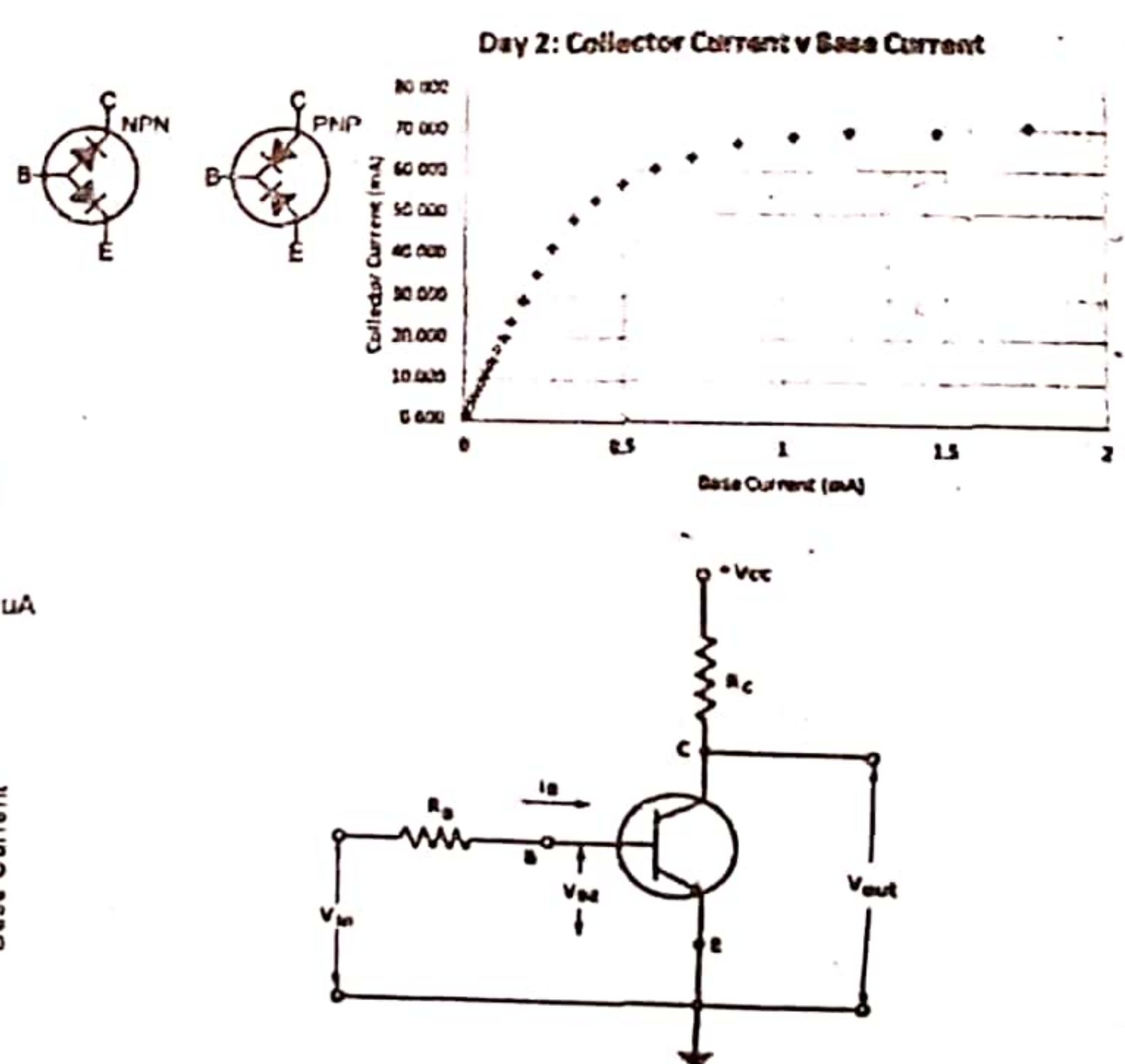
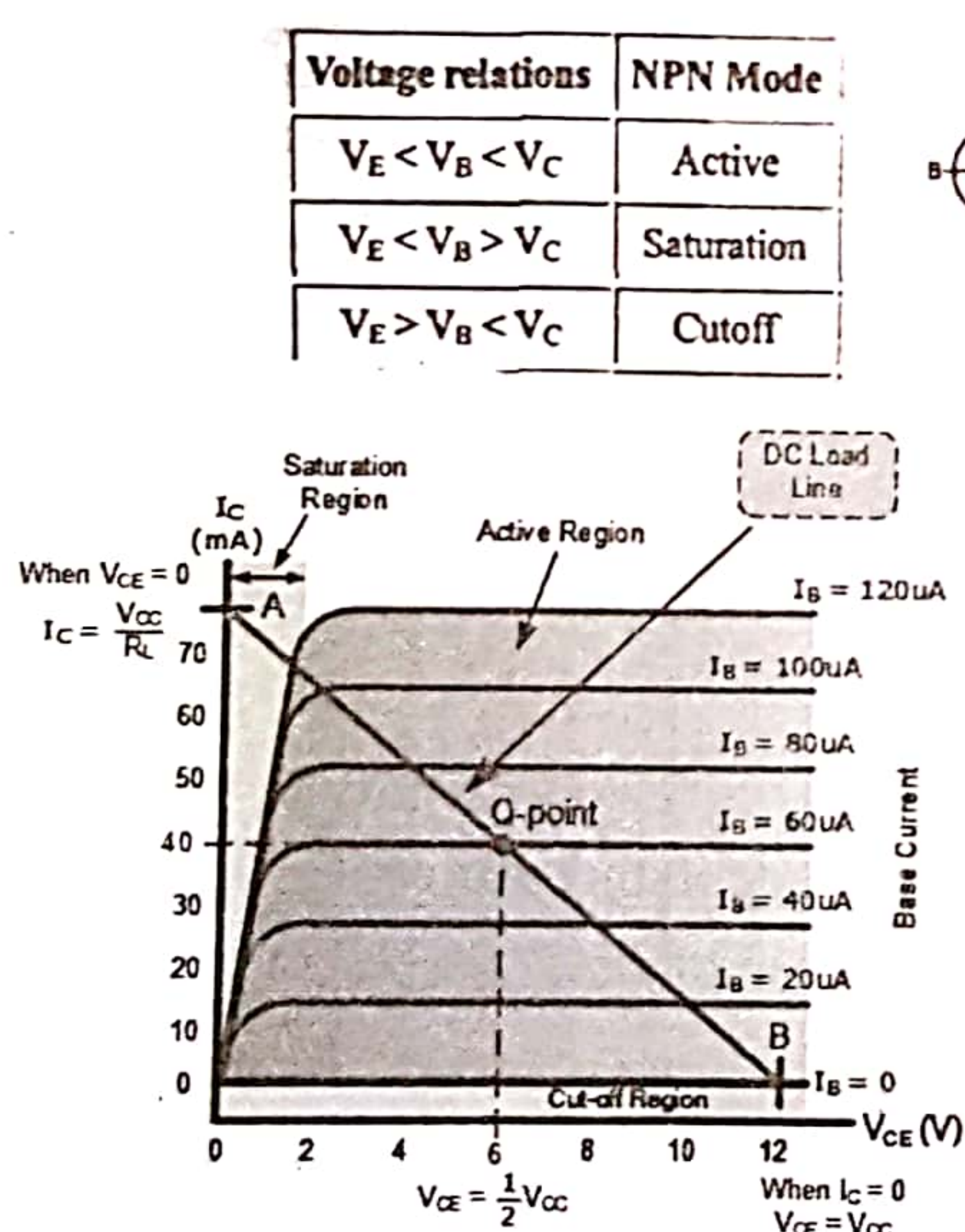
The dimensions of the constants have been asked in previous past papers many times. If you need the dimensions of G , then according to the equation $F = Gm_1m_2/r^2$, the dimensions of $G = (\text{force} \times \text{square of the distance})/\text{square of the mass} = MLT^{-2} \cdot L^2/M^2 = L^3M^{-1}T^{-2}$

3. When a bipolar junction transistor operates in saturation mode, a further increase in base current
 (1) turn on the transistor. (2) turn off the transistor.
 (3) increase the collector current. (4) decrease the collector current.
 (5) not change the collector current.

09

Transistors

When a npn transistor is working on its saturated mode, both the base - emitter and the base -collector junctions are in forward biased mode. Then the transistor is working as a closed switch and the value of I_C and I_E values are at their maximum. According to the meaning of the word 'saturated', you can understand that the current (I_C) attains to a maximum value. However, once something is saturated, it cannot increase anymore. Therefore, when I_C is saturated to a maximum value, it will not get increased when I_B is increased. When the transistor is in the saturated state, it is in the closed situation. Therefore, it remains as it is. If you think of relations, then $V_{CC} = I_C R_C + V_{CE}$ (look at the circuit). When the transistor is saturated, its V_{CE} value gets very small (0.2 V for a Silicon transistor). Therefore, $I_C = V_{CC}/R_C$. This is the maximum value that I_C can take.



4. According to the evidences found in Particle Physics, matter is composed of
 (1) 6 quarks. (2) 6 leptons.
 (3) 4 quarks and 4 leptons. (4) 6 quarks and 4 leptons.
 (5) 6 quarks and 6 leptons.

11

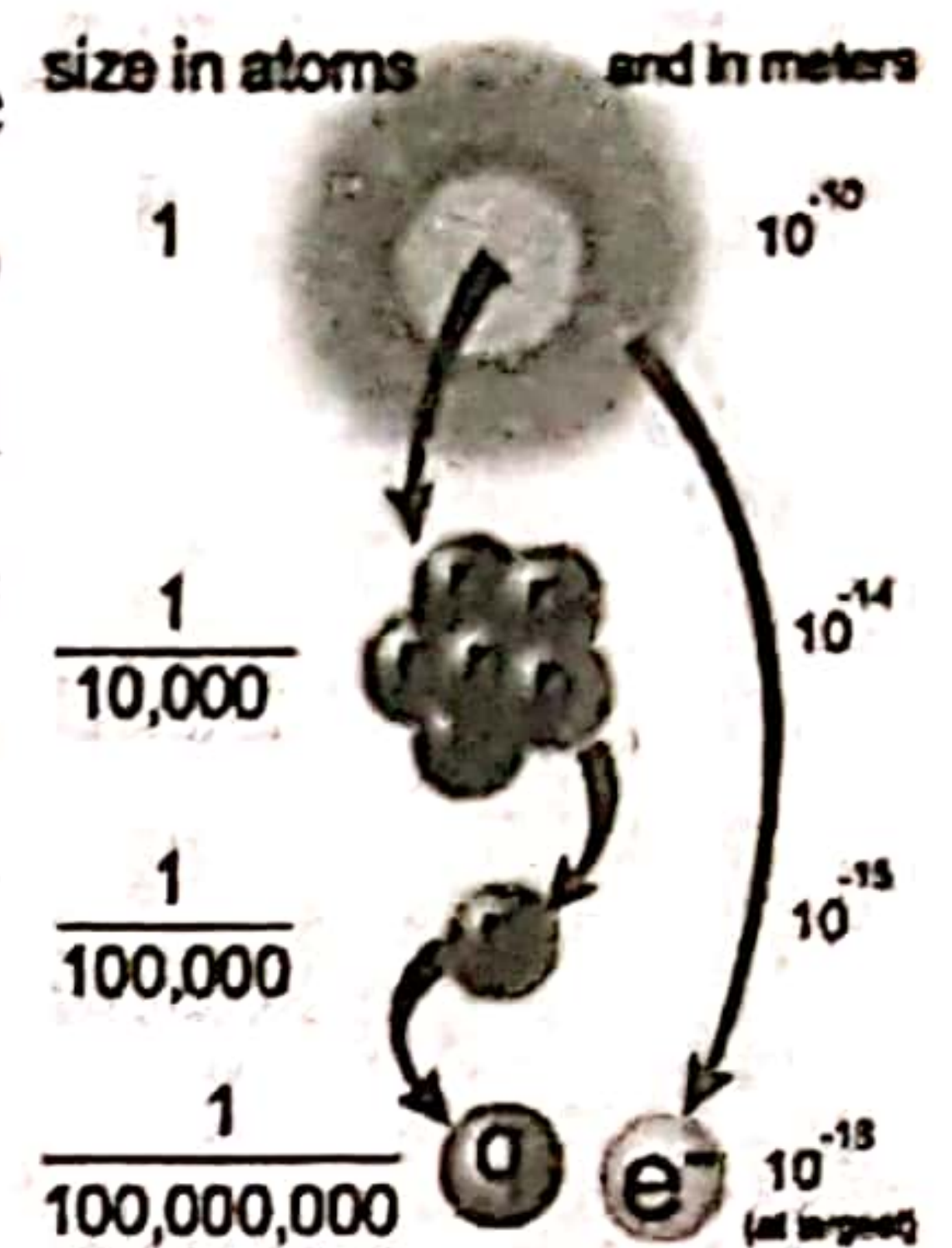
Partial Physics

From ancient times, the philosophers have thought about the basic elements of different components that the world/universe have been made. According to Aristotle's analysis, earth, air, water and fire have been considered as the basic elements. We need the earth to set foot on the ground. We need air to breathe. We need water to drink. We need fire to cook foods but not to give sorrow to others. Even Lord Buddha has

preached the things around us as Patawi, Aapo, Thejo and Waayo. The philosophers of China added metal and wood whereas the philosophers of India added space in addition to the elements. In our national flag, the four fingers of the lion depict earth, water, air and fire.

According to science, everything in the universe is made from a group of molecules. The molecules are made from atoms. In early days, the scientists had a view that the atom cannot be undivided anymore.

The word 'atom' is broken from the Greek word 'atomon'. Its meaning is undividable anymore. In Sinhala, the word 'පරමාණුව' is based on joining the two words of පරම and අණුව. පරම means there is nothing beyond it. When we say පරම සුවය (eternal comfort), පරම සනුට (eternal happiness), පරම ආදරය (eternal love), they indicate the maximum experience that you can get. Some get eternal comfort when they take some alcohol. When eternal comfort is at other places for someone like the office or the places where they go, there is eternal sadness to stay at home.



But we know that the atom is made from nucleus and electrons. The internal composition of electrons has not been found yet. Therefore, the electron is considered as a fundamental particle. But it has been found that the constituents of the nucleus that are protons (positive) and neutrons (neutral) are not fundamental particles. All the evidences have obtained to prove that they are made from a collection of very small fundamental particles called quark.

According to this, the constituents of the matter are considered to be belonged to the lepton family (where electron belongs) and the quark family. The members of these two families are shown here. The meaning of lepton is light weight. But in lepton family τ (tau lepton) is greater in mass. Before the knowledge about quark, the particles found by the particle physicists belong to three families. According to that, there are lepton, meson (with middle mass) and baryons (with high mass). It was found that meson is made from the combination of a quark and anti-quark whereas the baryon is made from three quarks. So, the scientists identified that three families can be reduced to two. But it was agreed to keep the family of lepton as it is.



There are six members in each of lepton and quark family. Electron has an elder sister which is with higher mass as muon (μ) and another elder sister with more mass as tau lepton (τ). The letter μ is in Greek alphabet to the middle whereas τ is to the end. The charge of e^- , μ^- and τ^- is $-1.6 \times 10^{-19} \text{ C}$ (that means the electronic charge). All these three have neutrino companions without charge as ν_e , ν_μ , ν_τ (neu e, neu mu, new tau). The meaning of neutrino is invisible. As there is no charge in neutrino, they could not be directly discovered by experiments. The name was given to these fundamental particles due to that reason.

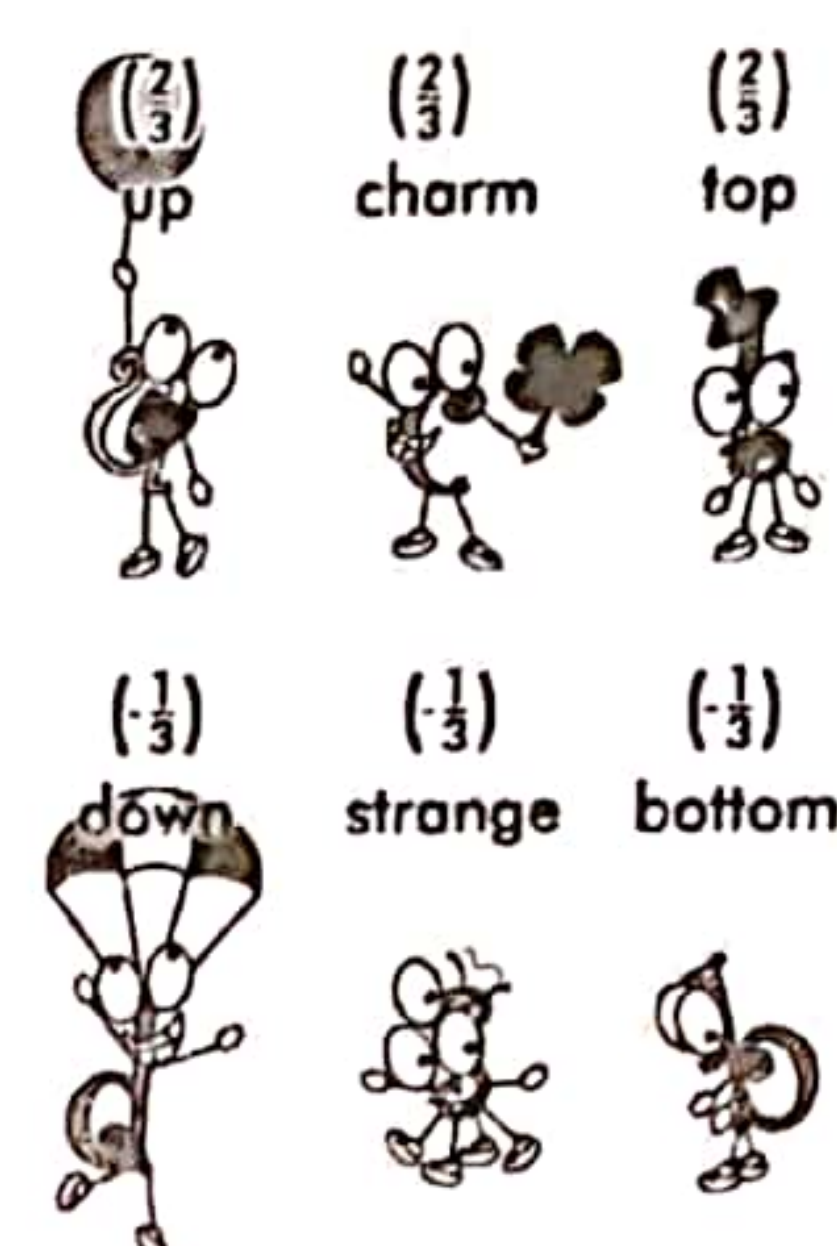
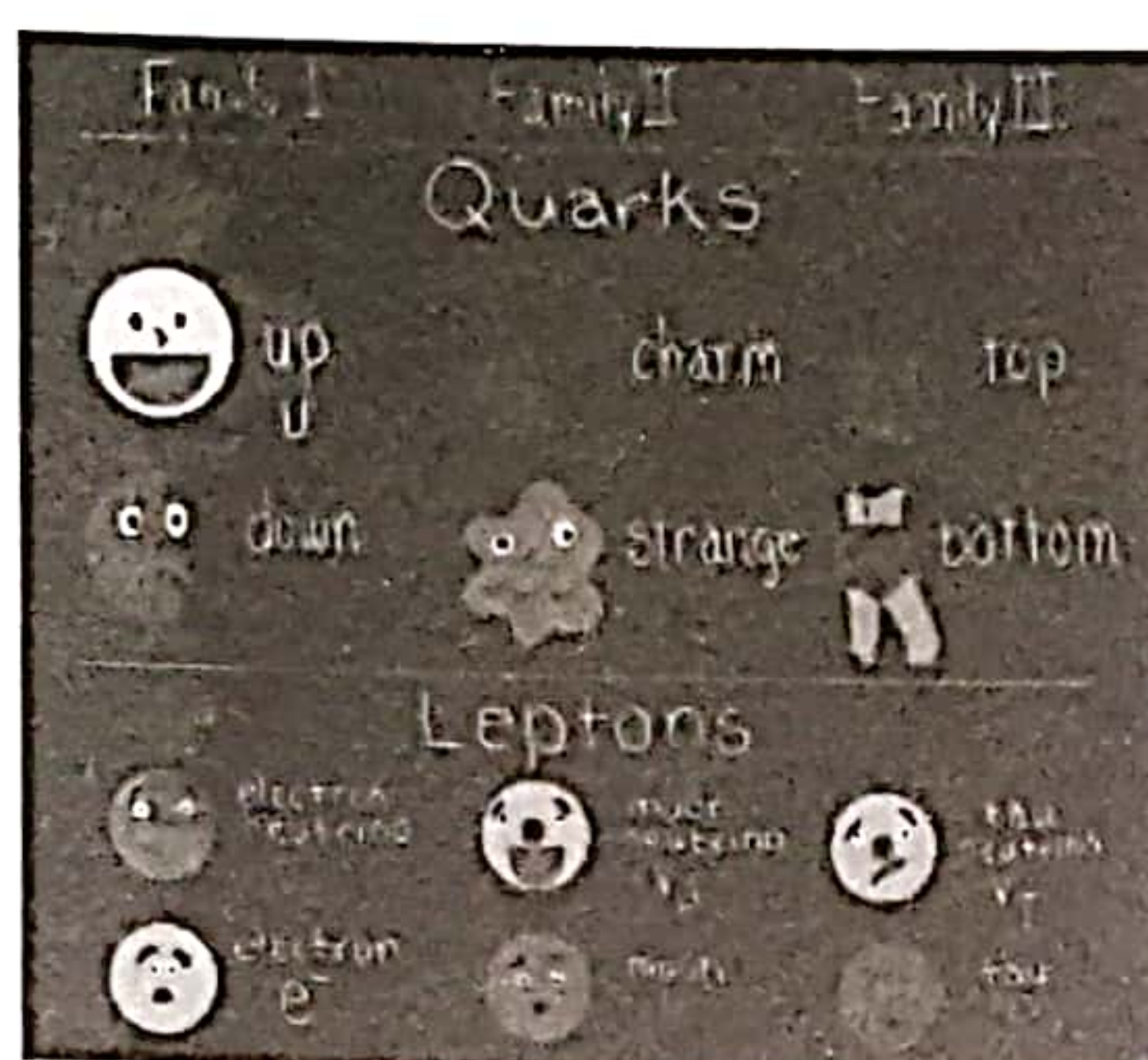


The scientists have given the names of u (up), d (down), c (charm), s (strange), t (top), b (bottom) for quarks. There is a reason to name like this way. Do not turn these names into Sinhala. It is better to be called as they are. For each quarks it is called as their flavors. Commonly a quark is represented by the symbol of q . Normally we say that a quark is with six flavors.

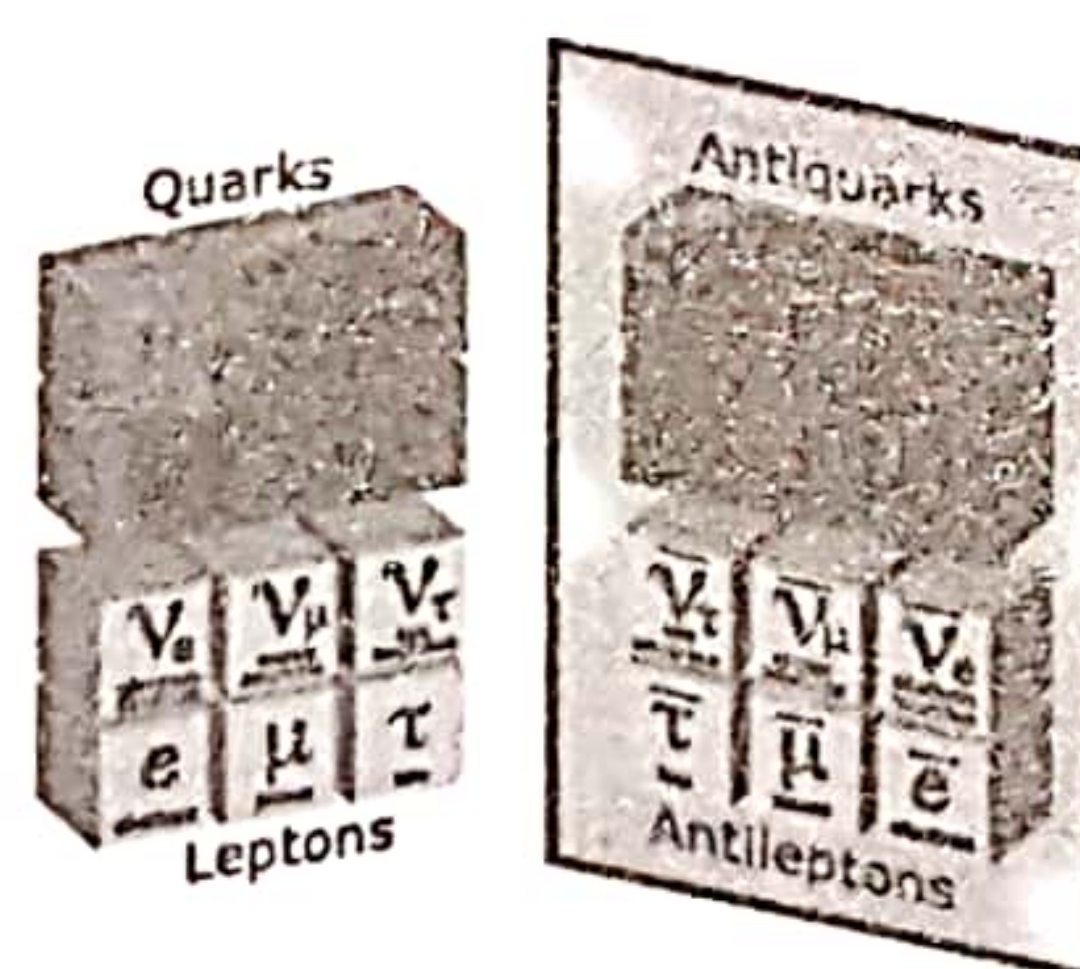
The first column of e^- , ν_e and u , d quark strips are known as the first generation where as the second column of fundamental particles are known as the second generation. Finally, the last column of particles are considered to be belong to the third generation. The charge of quarks are taking fractional values. Even though the charge of e^- , μ^- and τ^- is $-e$, the charge of quarks u , c and t are $+2/3 e$ while the charges of d , s

and b quarks are $-1/3 e$. Do not forget when writing quarks u , c and t , they should be written in the first row whereas when writing d , s and b quarks, they should be written in the second row. If you want to remember about u , c and t then remember the phrase 'university of Colombo teacher'. As quarks can be added to make other known particles (except leptons) and each charge of a particle is taking an integer value of an electronic charge such as 0, the charge of quark should take fractional values.

	up	charm	top	gluon	Higgs boson
QUARKS	down	strange	bottom	photon	
	electron	muon	tau	Z boson	
LEPTONS	electron neutrino	muon neutrino	tau neutrino	W boson	GAUGE BOSONS



The artistic representations of quarks are shown here. An up quark is shown as a Helium balloon. That is because the balloon of Helium goes upwards. Down quark is like a parachute as parachutes always go downwards. Quark c is like a charming daughter who always look at her reflection in a plane mirror. Quark s is like a strange figure with many heads and legs. Quark t is like a gentleman with a hat on his head. Quark b is like a small boy who goes to school by bending forward with a bag. When I see this figure, I remember the bomber of Katuapitiya church. How much can you change the mind of a child by using an influencer?



The six leptons and six quarks that we considered have anti-particles belonging to each other. The complete set of the diagram is shown here.

The first anti-particle that the scientists found was positron (positive electron – e^+). Afterwards, for each particle, an anti-particle was found as if like husband and wife. For example, $e^- - e^+$; $\mu^- - \mu^+$; $\tau^- - \tau^+$; $\nu_e^- - \nu_e^+$. As a neutrino does not have a charge, its anti-particles are represented by applying a bar sign above ν_e , ν_μ , ν_τ . They are called as neu e bar ($\bar{\nu}_e$), neu mu bar ($\bar{\nu}_\mu$), and neu tau bar, ($\bar{\nu}_\tau$). If you consider mathematically, then by applying a bar sign, an inverse variable is represented ($x - \bar{x}$).

Likewise, the anti-quark of u is represented as \bar{u} and for every other anti-quark particle, we use a bar sign above each of them. So, we can decide that the universe has total 24 fundamental elements like up bar quark - \bar{u} ; down bar quark - \bar{d} . It consists of 6 leptons, 6 quarks, 6 anti-leptons and 6 anti-quarks. I consider this table as the periodic table of particle physicists.

Here are the laws of making particles by using quarks.

- (1) From any three quarks you can make all the other baryons like n, p . When all of these quarks are transformed into anti-quarks, you will get anti-baryons like \bar{n}, \bar{p} .
- (2) The addition of any quark and anti-quark creates mesons (pion - π^0, π, π^+).

Apart from this, any particle cannot be produced with any composition. The reason for this very simple. As mentioned earlier, the charges of every discovered particle should take a value like $0, \pm e, +2e$. The charge of all these particles cannot take a fractional value. You can decide that a zero or a whole number ($0, \pm 1 \dots$) can be obtained like the charge of an electron only from limited combinations that was mentioned above.

The valid combinations of quarks and the names of relevant particles (symbols) have been shown here. Apart from p, n there is no need to know more in A/L.

$$p = uud = \frac{2}{3} + \frac{2}{3} - \frac{1}{3} = +1 \quad \bar{p} = \overline{uud} = -\frac{2}{3} - \frac{2}{3} + \frac{1}{3} = -1$$

$$n = udd = \frac{2}{3} - \frac{1}{3} - \frac{1}{3} = 0 \quad \bar{n} = \overline{udd} = -\frac{2}{3} + \frac{1}{3} + \frac{1}{3} = 0$$

$$\Omega = sss = -\frac{1}{3} - \frac{1}{3} - \frac{1}{3} = -1 \quad \pi^- = \bar{u}d = -\frac{2}{3} - \frac{1}{3} = -1$$

$$\pi^+ = u\bar{d} = +\frac{2}{3} + \frac{1}{3} = +1 \quad \pi^0 = u\bar{u} \text{ or } d\bar{d} = \left(+\frac{2}{3} - \frac{2}{3}\right) \text{ or } \left(\frac{1}{3} - \frac{1}{3}\right) = 0$$

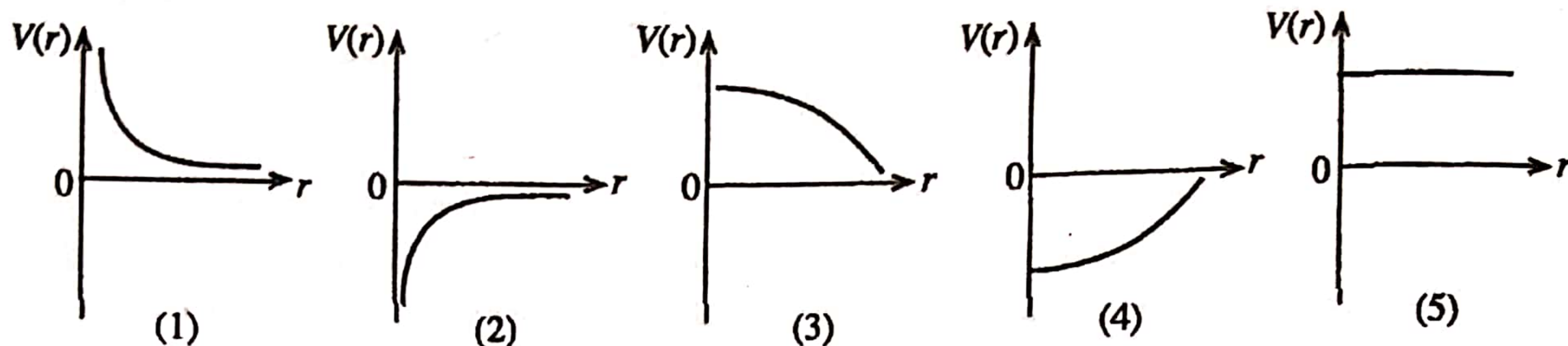
$$\Delta^{++} = uuu = \frac{2}{3} + \frac{2}{3} + \frac{2}{3} = +2 \quad \Delta^- = ddd = -\frac{1}{3} - \frac{1}{3} - \frac{1}{3} = -1$$

Putting names to quarks can be understood by this way. The scientists consider p and n as two family members in a family. There is a net charge for proton where the net charge is zero for neutron. Therefore, p , proton is considered as the elder brother and n is considered as the younger sister. As there is a charge, proton is considered to be in a higher place. According to that in the family of p and n , p is considered as the up companion whereas n , is called as the down companion. So as two up quarks are needed to make a proton, that is called as an up quark. As two down quarks are needed to make n , it is called as a down quark. During the middle of 20th century, particle physicists have discovered many particles (apart from p, n). It has become a challenge to explain that why there are there. Therefore, such particles that cannot be explained were called as strange particles. To build them, they need another quark other than u and d . That quark is known as strange.

The name 'charm' came like this way. When scientists who study about quarks gathered at an instance, the sentence of "What a charming theory is this" was said. Accordingly, the next quark was named as charm. The last two quarks of top and bottom were name for the convenience in the vocabulary usage. How much do they match if we use up and down or top and bottom? Initially, they were named as truth and beauty. These two characteristics are essential to human beings. But later on, that was abandoned and everyone liked to use as top and bottom.

This model has been named as the 'standard model of the universe'. By using this model today, not only particle physics, the evolution of the universe as well as many things related astronomy, cosmology subjects can be explained. The answer for the question is 6 leptons and 6 quarks. The question asked about how the matter has been made.

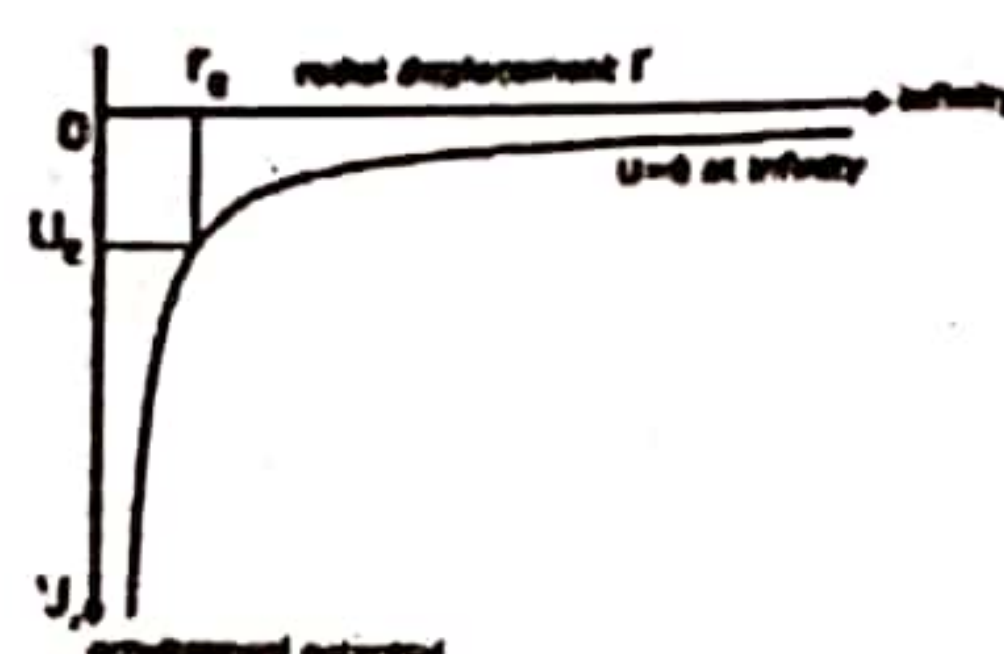
05. The variation of the gravitational potential $V(r)$ due to a point mass, with distance r is best represented by



05

Gravitational Force

The gravitational potential (V) of a point mass of m with a distance of r can be given by $V = -GM/r$. Therefore, the variation of V with r can be represented simply as shown below.



The negative sign in the expression is very important. As we know, gravitational force is an attractive force. Therefore, two objects with masses are bonded with each other due to gravity. When the force is attractive, the potential takes a negative value. The potential at a distance r_1 $V_1 = -GM/r_1$, at a distance r_2 (let us take as $r_2 > r_1$), $V_2 = -GM/r_2$. As $r_2 > r_1$, $V_2 > V_1$. This indicates that we need to do some work if we need to take away and separate one mass from another mass. If it is a repulsive force, then we do not have to do work for the separation. We need to do bad activities if we need to break the relation between two people. You need to defame each other as much as possible. When $r \rightarrow \infty$, then $V \rightarrow 0$. However, at infinity (when they are separate as much as possible), it is the normal tradition to take the gravitational potential as zero. But when $r \rightarrow 0$, then . This seems bit non-physical. This happens as we consider a point mass here. If we consider a solid sphere with radius R , then the gravitational potential inside the sphere with a distance r from the centre is $V(r) = -\frac{GM}{2R^3}(3R^2 - r^2)$ [this is not in the syllabus]

When $r = R$, we get the expression we know which is $V(r) = -GM/r$.

When $r = 0$, $V(r) = -3/2 \cdot GM/R$. This will not get infinite.

06. Which of the following statements is incorrect regarding thermometry?

- (1) There must be a measurable physical quantity that varies with temperature.
- (2) Mercury-glass thermometers consist of thin-walled glass bulbs.
- (3) By using a mercury-glass thermometer with a large mercury bulb, the range of measurements can be increased.
- (4) Two different types of thermometers may give slightly different readings at the same temperature as all thermometric properties are not equally sensitive.
- (5) Having a large contact angle between mercury and glass is an advantage for accurate readings in mercury-glass thermometer.

04

Thermometry

There should be a measurable characteristic for a thermometric material which varies with temperature. Otherwise, what should be measured? This has been asked in many past papers. Therefore, it is not essential for the measuring characteristic (quantity) to have a linear variation with the temperature. It is better if it can vary linearly.

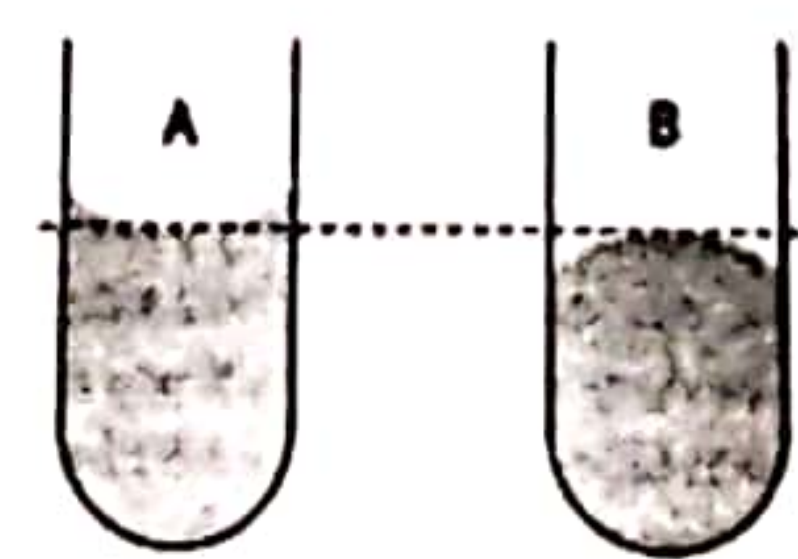
In a mercury-glass thermometer, the bulb is made from glass with thin walls. Then the thermometer quickly responds. The heat is quickly transferred towards mercury. So that the temperature difference between the mercury in the bulb and the liquid in which the temperature is being measured will be less. It is negligibly small. The measured reading from it is getting accurate too. The temperature range which can be measured cannot be changed by enlarging the bulb. Normally, the temperature range which can be measured from mercury is from -38°C (melting point of mercury) to 356°C (boiling point of mercury). This range belongs to mercury. Therefore, without changing the thermometric liquid, generally you cannot change the range of temperature that can be measured. By having a bigger bulb can occupy a bigger volume of mercury and can make the thermometer more sensitive. It gives a wide expansion for a certain change of temperature but the given reading may not be very accurate due to the absorption of more heat.

When taking the readings from two different type of thermometers, there can be a little change in between the readings from each of them. There can be many reasons for this. The method in which the thermometric material gets the heat can be different. The materials may not have the same sensitivity. The accuracy can be changed due to the absorbed heat amount. When the thermometric material changes, the characteristic that is sensitive to the temperature obviously changes.

The surface tension of mercury is higher. Therefore, the contact angle it makes with glass is greater. Actually, the special characteristic that is relevant to this is that mercury does not wet the glass. Therefore, the readings can be done accurately. As the surface tension of it is higher, it does not go and stick with others.

The first mercury-glass thermometer is made in 1714 by Daniel Gabriel Fahrenheit.

Why does the Fahrenheit scale start from 32? Fahrenheit thought the body temperature as 96°F . If so, then the melting point of water was taken as 32°F because the difference between 96 and 32 is 64 and 64 simply divides from 2. Then calibration is easy. Next, according to this scale, the boiling point of water was obtained as 212. The difference between 32 and 212 is obtained as 180. Fahrenheit loved this difference. The temperature difference between the melting point and the boiling point is 180. That means there is a sort of opposite characteristic in it.



07. Consider the following statements regarding the physical properties of ultraviolet and ultrasound waves.

- (A) Energy of both waves depends on their frequencies.
- (B) Both waves have the ability of ionizing materials.
- (C) Both waves can be polarized.

Which of the above statements is/are incorrect?

- (1) Only A
- (2) Only A and B
- (3) Only A and C
- (4) Only B and C
- (5) All A, B, and C

Electromagnetic Waves

13

Does the energy of an electromagnetic wave and a mechanical wave depend upon their frequency? To me, this is a bit debatable question. Once you read this and if you do not agree with me, then please let me know. Ultraviolet light is a magnetic wave. The ultrasound waves are high frequency sound waves ($> 20 \text{ kHz}$) that are not sensitive to our ear. The expression for the energy of a sound wave is not in the syllabus. But I

will try to get the related expressions to this in a simpler way for your understanding. How can you find the energy of a sound wave? It can be simply interpreted as the energy or the intensity (Wm^{-2}) across a unit area in a unit time.

$$\text{Intensity} = \text{energy} / (\text{time} \times \text{area}) = (\text{energy} / \text{time}) \times (\text{distance} / \text{volume}) = (\text{energy} / \text{volume}) \times (\text{distance} / \text{time}) \\ = (\text{energy} / \text{volume}) \times \text{the wave speed}$$

The energy of the wave is obtained from the simple harmonic motion of the particles. Therefore, the total energy is equal to the maximum kinetic energy.

$$\text{Energy} = \frac{1}{2} m v_{\text{max}}^2 = \frac{1}{2} m (A\omega)^2 \text{ where } A - \text{the amplitude of the wave, } \omega - \text{angular frequency of the wave} \\ (\text{energy} / \text{volume}) = \frac{1}{2} (m / \text{volume}) (A\omega)^2 = \frac{1}{2} \rho (A\omega)^2 \text{ where } \rho - \text{the density of the medium}$$

$$\text{Therefore, intensity} = \frac{1}{2} \rho A^2 \omega^2 v = \frac{1}{2} \rho v A^2 \omega^2 \text{ where } v - \text{the speed of the wave}$$

According to this, the intensity of the wave is dependent upon $\omega^2 / (2\pi f)^2$. Likewise, it is proportional to the square of the displacement amplitude (A^2). But a sound wave is felt by us as a pressure wave. The sensation that the ear feels is known as loudness. The above expression can be modified like this way.

$$\text{Therefore, intensity} = \frac{1}{2} (\rho v A \omega)^2 / \rho v. \text{ We will consider the units of } \rho v A \omega.$$

$$\rho v A \omega = \text{kg/m}^3 \cdot \text{ms}^{-1} \cdot \text{m} \cdot \text{s}^{-1} = \text{kgms}^{-2} / \text{m}^2 = \text{Nm}^{-2}$$

There are dimensions of pressure for this. Therefore, the term $\rho v A \omega$ can be identified as the pressure amplitude (P_0). Accordingly, the sound intensity can be expressed as $\frac{1}{2} P_0^2 / \rho v$. Our ear detects pressure variations not displacement variations. The ear is a pressure sensor. According to this, the ear has sensations of pressure amplitudes not displacement amplitudes. Therefore, the loudness is proportional to the square of the pressure amplitude of the sound wave where loudness does not depend on the frequency. It is true.

The ultrasonic waves are not heard by our ear. Therefore, when talking about ultrasonic waves we cannot talk about loudness. The detectors (piezo-electric crystals) which identify ultrasonic waves detect them as pressure waves. If you consider like this way, then you can say that the intensity of ultrasonic waves is not dependent upon frequency. But there is $\omega / 2\pi f$ in P_0 . If you look from that point, then the intensity is dependent upon the frequency. This is a very complex problem to me as I cannot give a straight forward answer. Even, children do not know any of these equations.

According to the syllabus, the children know that the loudness of a sound wave is proportional to the square of the pressure amplitude only. They do not know about the energy (intensity) of an ultrasonic wave. If somebody can clarify this matter, then please let me know.

When an electromagnetic wave is considered as a wave, the children do not know the expression for its intensity as it is not in the syllabus. But I will try to derive an expression for the intensity in a simple way.

Consider a parallel plate capacitor with plate area A and the distance between plates d . If the potential difference between the plates is V , then the stored electrical energy of the capacitor is $\frac{1}{2} CV^2$ where C is the capacitance of the capacitor. This energy can be considered to be stored in the area where the electric field is applied between the plates. Therefore, the stored electrical energy (energy density) in a unit volume in that region can be interpreted as $\frac{1}{2} CV^2 / Ad$. $C = A\epsilon_0 / d$ and $V = Ed$. According to that the energy density = $\frac{1}{2} (A\epsilon_0 / d) (E^2 d^2 / Ad) = \frac{1}{2} \epsilon_0 E^2$

Even we have derived this expression for a parallel plate capacitor, for the energy of an electric field at a vacuum per a unit volume generally can also be deduced with the same expression.

Now for an electromagnetic wave also we can use this expression. As the electric field of an electromagnetic wave is varying with the time, there will also be a magnetic field corresponding to that variation. Therefore,

the energy is stored both in the electric field and in the magnetic field. I am not going to derive the expression for the stored energy in the magnetic field. If I do so, then the children of A/L standard will not be able to digest that.

But from a simple argument, you can get an expression for the energy density of electromagnetic wave. The stored energy of an electromagnetic wave is divided by 50% by each of the two fields. Both electric and magnetic fields are dependent upon each other. Both are equal partners of the wave. Therefore, the energy is not divided in a biased way, especially as more portion for one field and less portion for the other field. If you are equal partners, then you need to have equal rights.

Now we will consider a sinusoidal electromagnetic wave. If E_0 is its maximum electric field intensity (the maximum of the electric field- amplitude), then the mean square value of the electric field intensity is $E_0/\sqrt{2}$. Therefore, the stored energy density of the electromagnetic wave is $2 \times \frac{1}{2} \epsilon_0 E_0^2 / 2 = \frac{1}{2} \epsilon_0 E_0^2$

Actually, when we add the stored energy of the magnetic field and express in B and E, you will get the above expression.

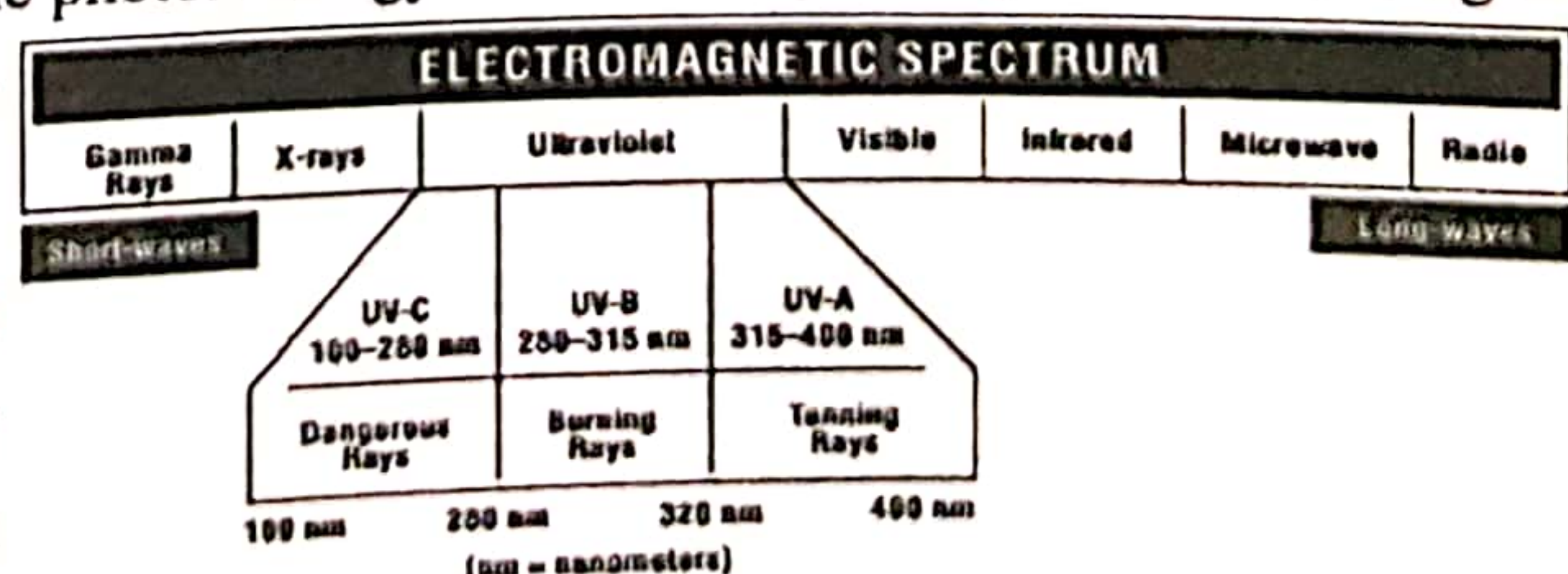
Now when energy density is multiplied by the speed of electromagnetic wave (c), you will get the intensity. $I = (\text{energy/volume}) \times \text{speed} = \frac{1}{2} \epsilon_0 c E_0^2$

According to this, the intensity of an electromagnetic wave does not depend upon the frequency due to wave model clearly. Therefore, according to the wave model, the energy of ultraviolet waves does not depend upon the frequency. Even if ultrasonic waves are treated as pressure waves, then we can consider that they are not dependent upon frequency as well. If we consider that ultrasonic waves are dependent upon frequency [$I = \frac{1}{2} \rho v A^2 (2\pi f)^2$], then the statement that mentions the energy of both waves are dependent upon their frequencies is false. But there is another story here. According to photon theory, the energy of electromagnetic waves is dependent upon its frequency. The photoelectric effect cannot be explained from the wave model of electromagnetic waves. This confusion was solved by Einstein by quantizing the electromagnetic waves and taking the energy of a photon as hf . Likewise, sound waves can be quantized. Sound waves as well as ultrasonic waves are produced by the vibrations of a medium or molecules. When the energy of sound waves is divided into energy packets, they are called as phonons. Their energy is also given by hf . According to this, the energy of ultraviolet waves as well as ultrasonic waves are dependent upon their frequency.

Especially, the ability of ionization for a wave is decided by the frequency of their waves. The incident photons or phonons should have an ability to remove electrons. When there is such a collision, that means when an interaction happens, the electron removal ability cannot be obtained from the wave model. When there is a collision, it can be explained from the collision of an energy packet. There is a common recognition that the energy of an incident photon should be greater than 10 eV to ionize the atoms/ molecules of a medium. To remove the electron in the Hydrogen atom (to ionize) 13.6 eV is needed. To ionize a water molecule, about 33 eV is needed.

As shown in the figure, all waves (photons) in UV waves range cannot ionize a medium. UV waves are divided into three as UV-A, UV-B and UV-C. The photon energy which has the ability to ionize belong to the waves of shorter wavelength which belong to UV-C range.

The phonons of ultrasonic waves cannot ionize material any way. If we consider the frequency as 20 kHz, then the energy of a phonon of an ultrasonic wave is $hf = 4.14 \times 10^{-15} \text{ (eVs)} \times 20 \times$



$10^3 = 8.3 \times 10^{-11}$ eV. This value is no longer near to 10 eV relatively.

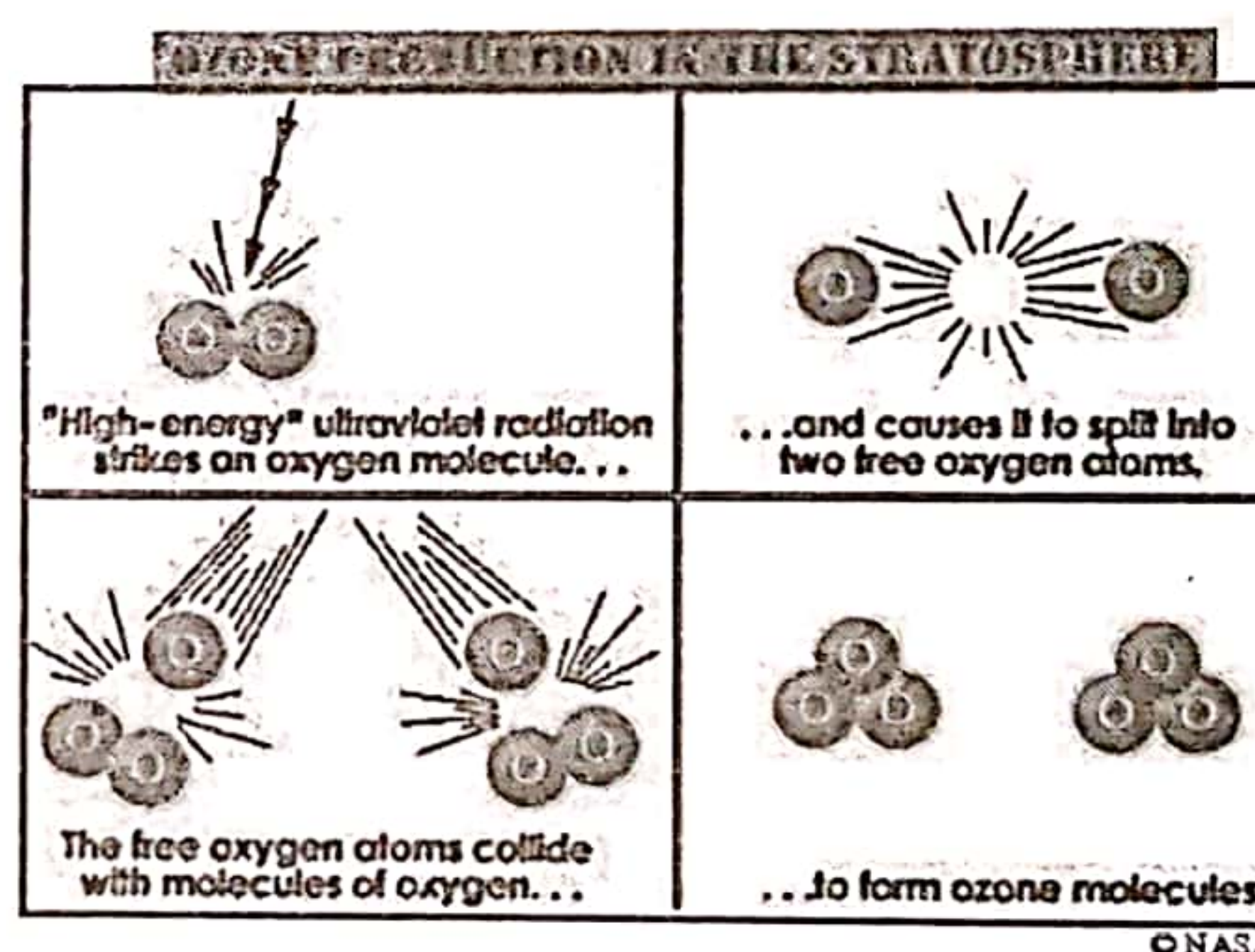
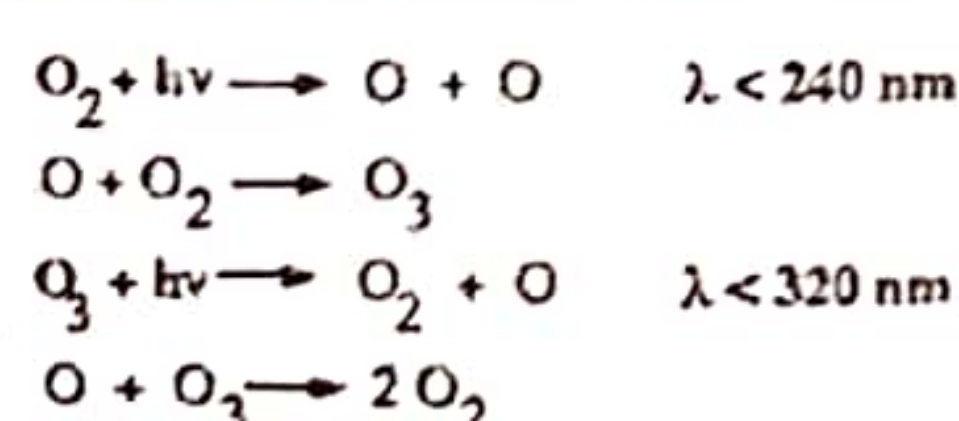
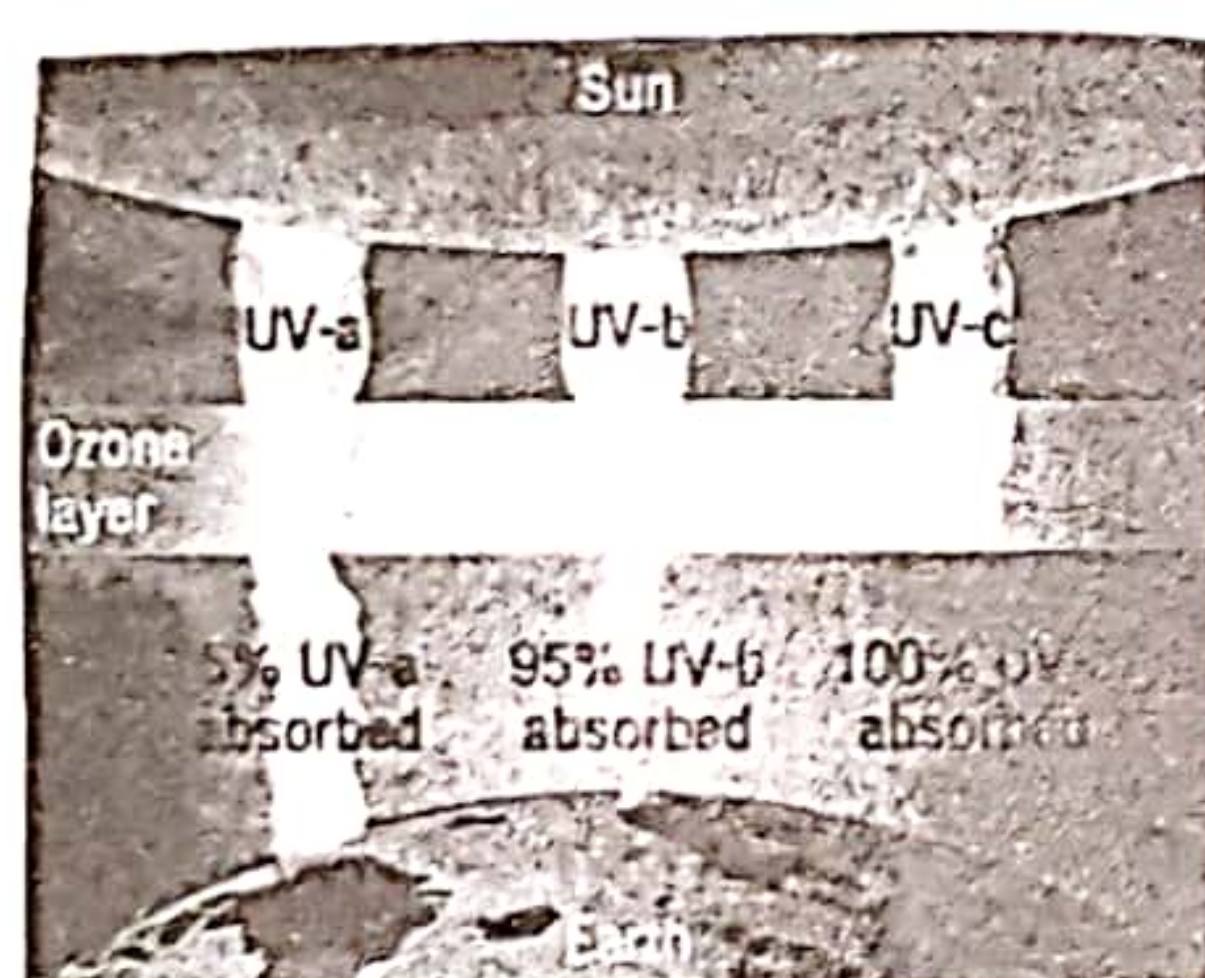
In air (speed = 330 ms^{-1}), the wavelength of a 20 kHz wave is 1.65 cm. In medical imaging and scanning work ultrasonics are being used due to this characteristic of inability of ionize. The statement of 'both ultraviolet and ultrasonic waves can ionize materials' is false. UV can ionize for a certain range but ultrasonics never contribute for ionization. So, the statement of both can ionize is a false statement. But the problem occurs for the first statement. To find the truth of the second statement, you must use frequency. The energy of both UV photon and phonon of an ultrasonic wave relies on the frequency.

Therefore, to check the accuracy of the first statement you need to use the wave model whereas to check the accuracy of the second statement you need to use the particle model. Therefore, there can be a conflict as I think.

In the third statement, there is no problem. UV waves are transverse waves (as they are electromagnetic waves). Ultrasonic waves are sound waves with high frequency. Sound waves are longitudinal. Only transverse waves can be polarized. So, both waves cannot be polarized. All this time, the word 'polarized' have been used. Now there is 'polarization' in the dictionary.

As it has not been used before, it can be an unfamiliar term. Polarization is not liquefaction.

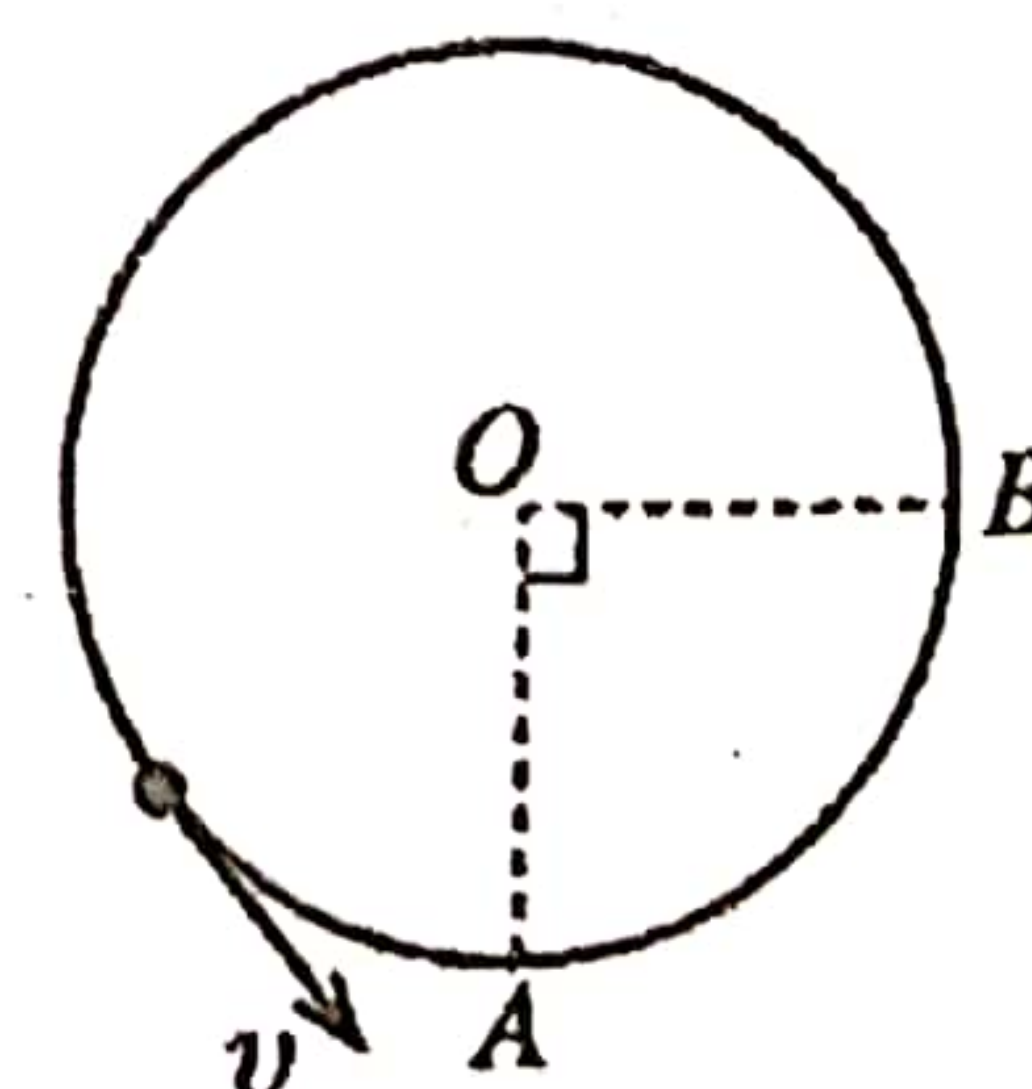
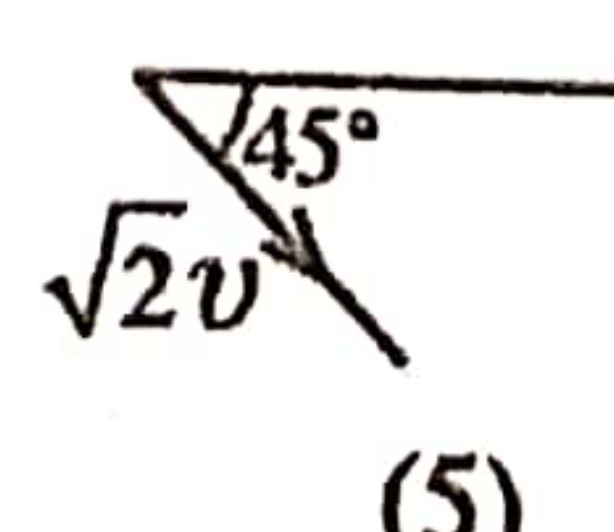
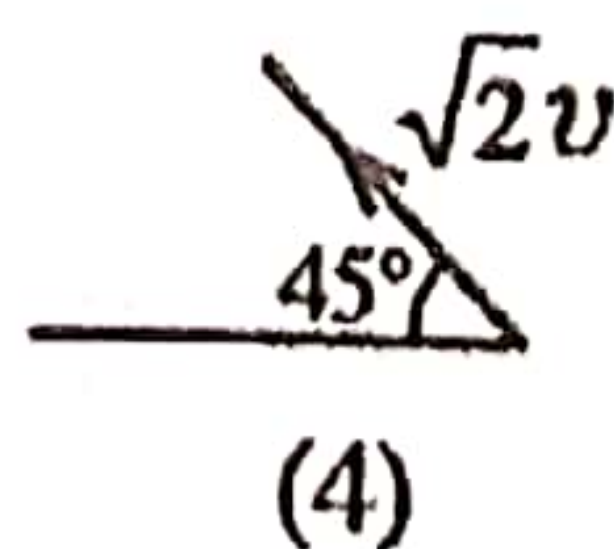
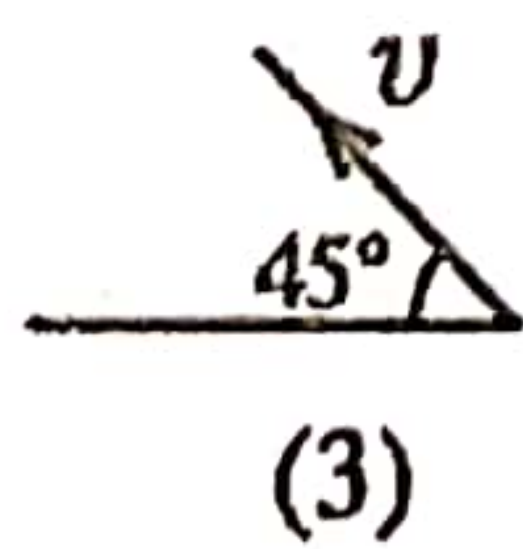
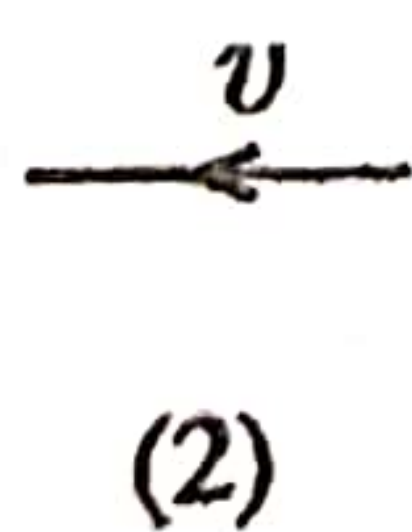
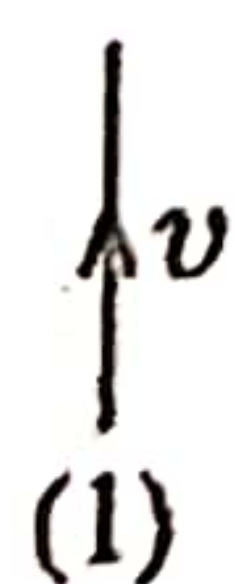
According to the shown figures, UV-C rays that belong to shorter wavelength range are more dangerous. The rays of UV-C range are being absorbed by ozone in the upper atmosphere. As mentioned earlier, UV ray photon of this range has a greater energy. They break O_2 molecule into two atoms, and make ozone molecule, O_3 with three oxygen molecules like this way.



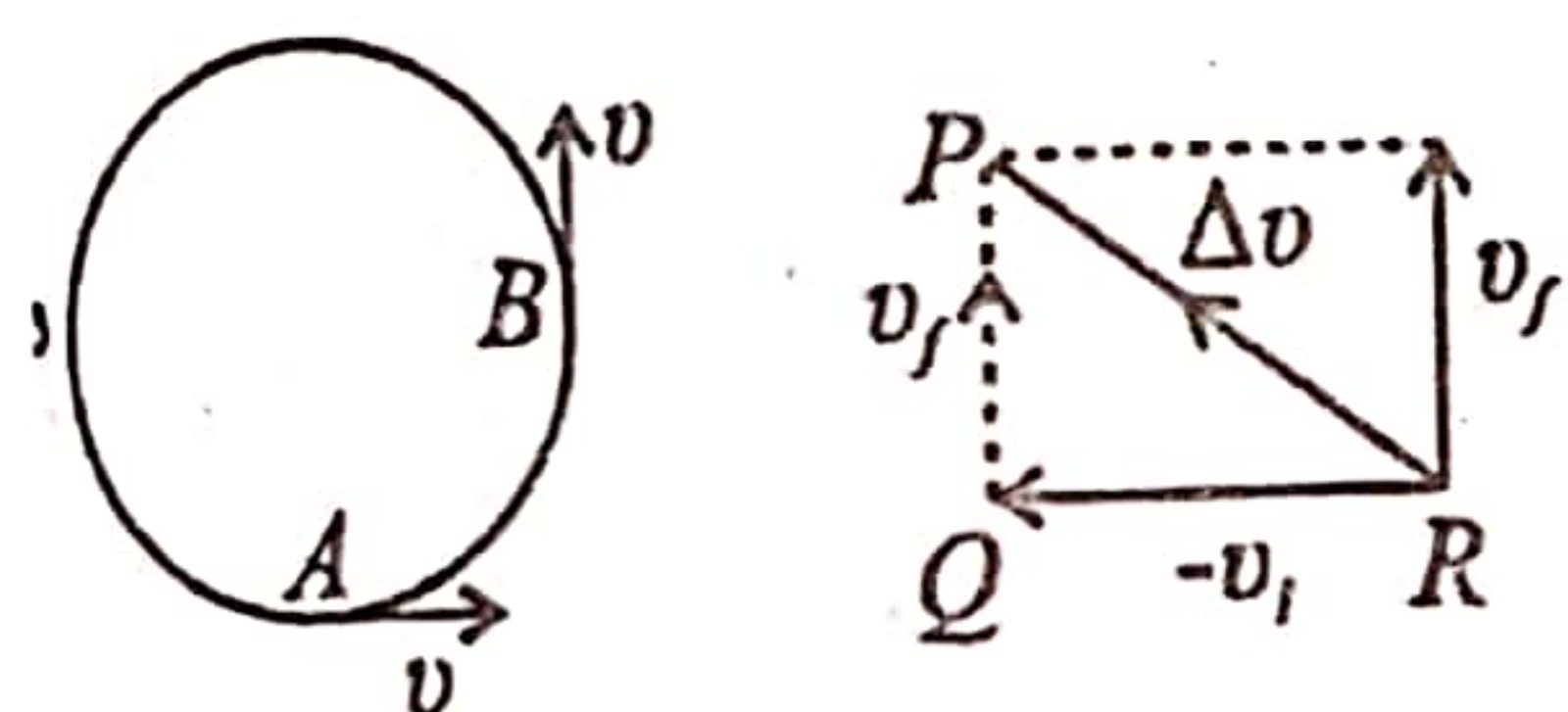
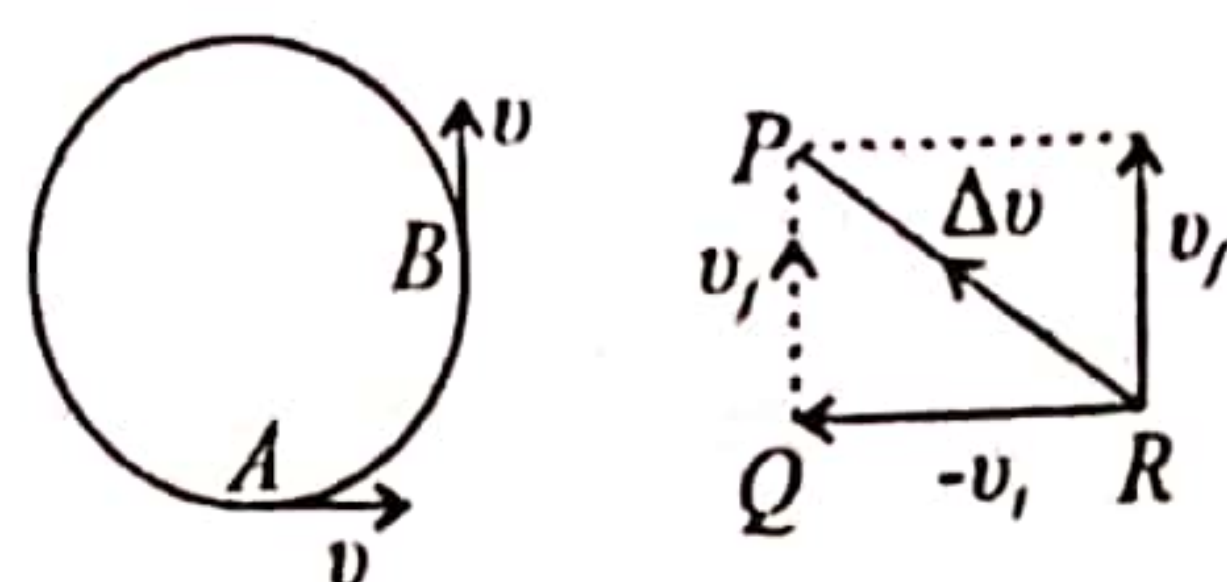
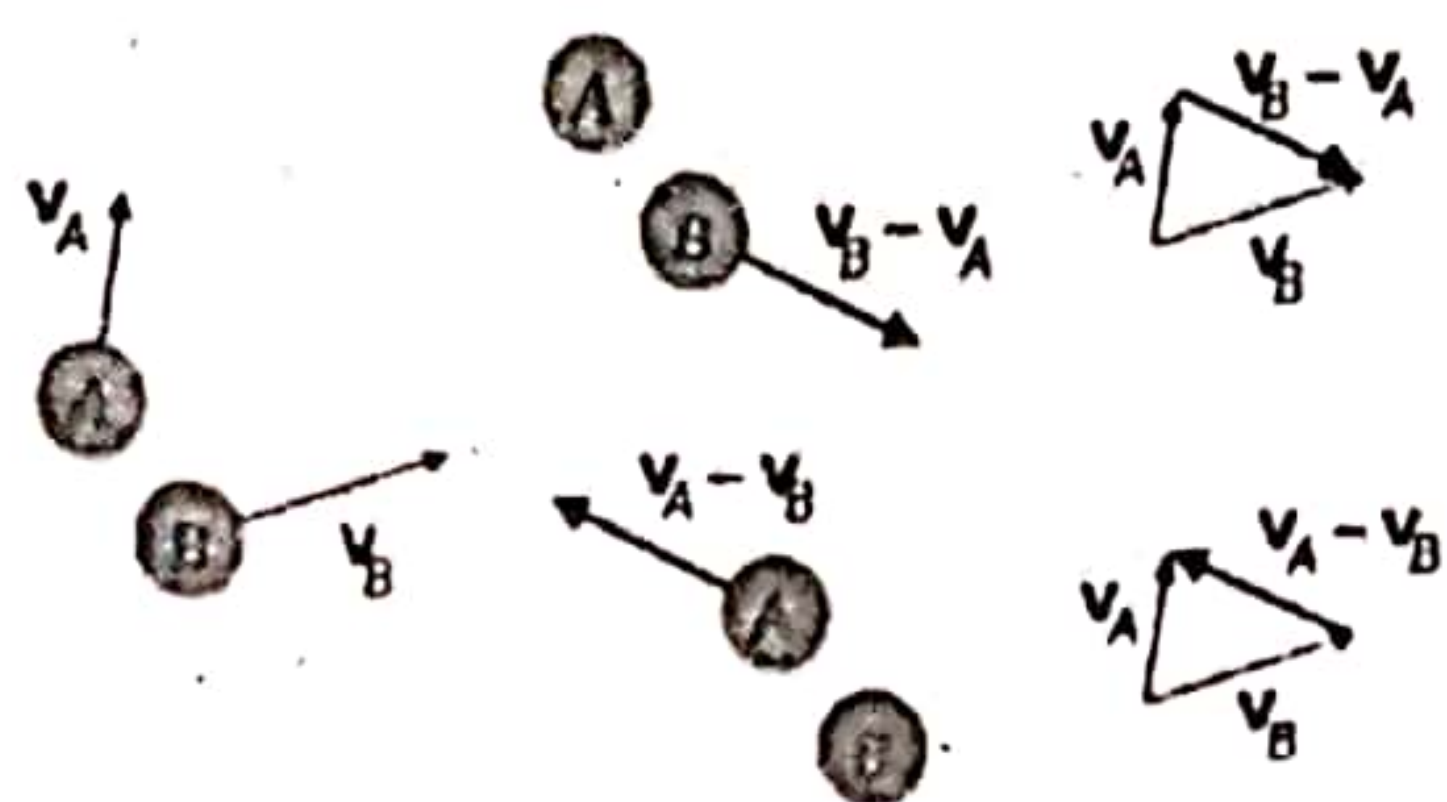
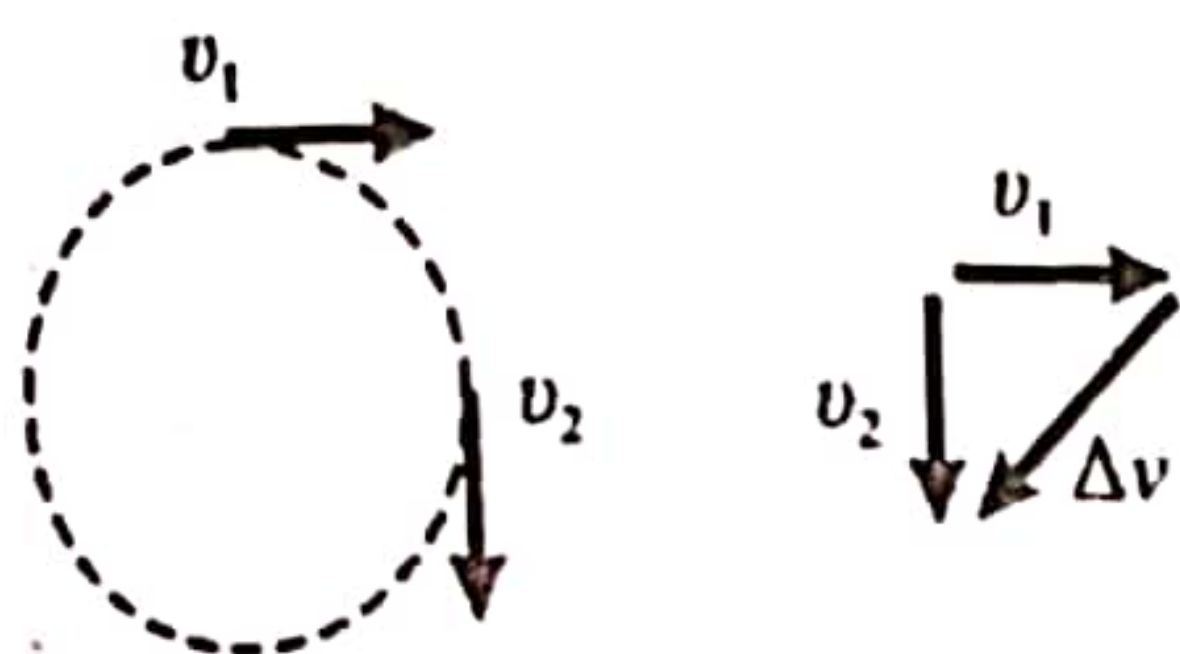
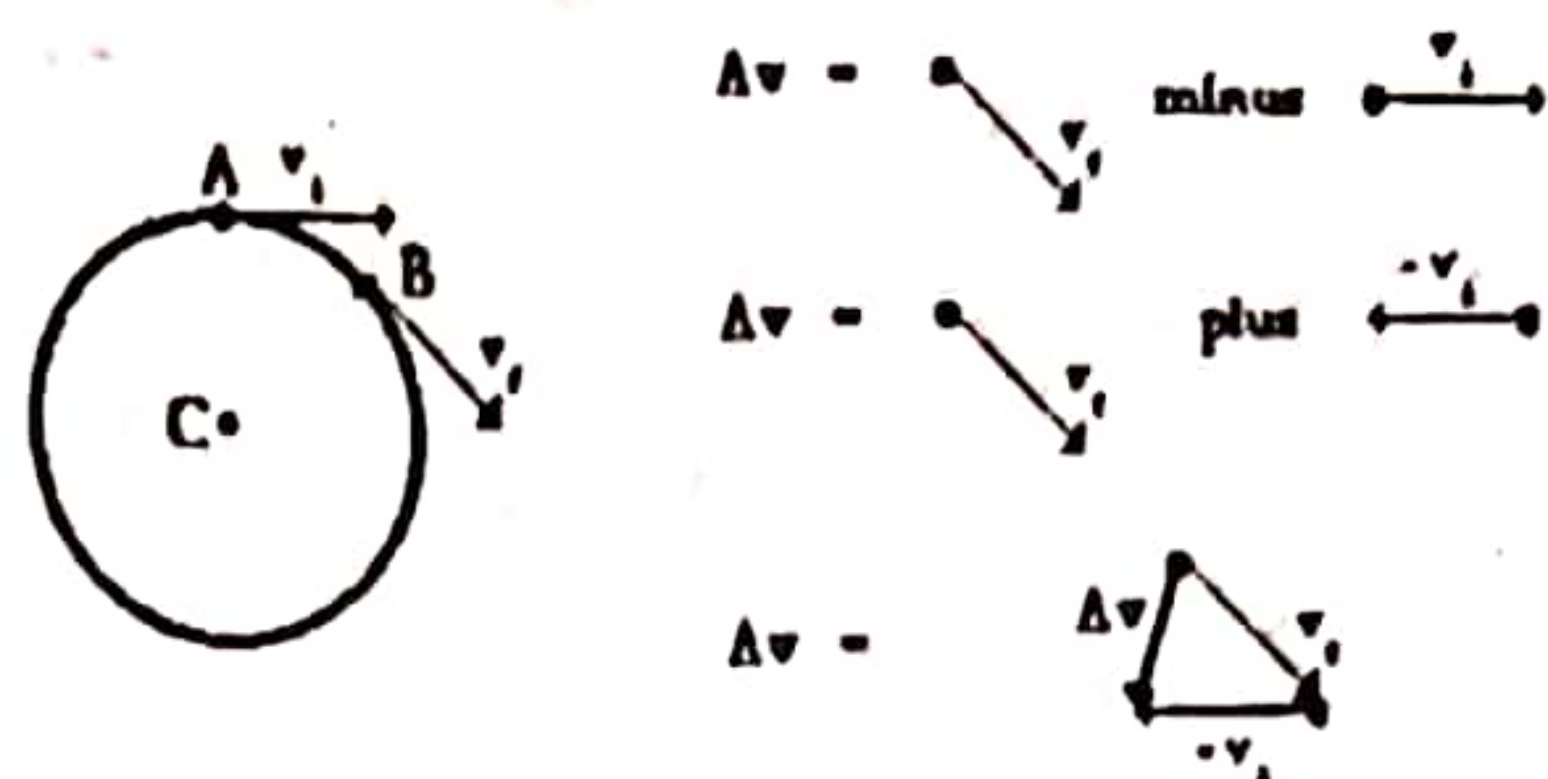
This can be done by UV photons with shorter wavelengths that are less than 240 nm. The reason is that the incident photon should have a higher energy than the binding energy of two oxygen atoms in O_2 molecule.

When UV photons that are less than 320 nm wavelength are hit with O_3 molecule, O_2 molecules are made. This process is a cyclic process. Again, O_3 is broken and O_2 is created.

08. An object is moving in a circular path at a constant speed as shown in the figure. The change in velocity of the object, when moving from A to B is



The figures show how to find the difference of two vectors that are directed to different directions. Likewise, to find the velocity difference of an object in a circular motion when it goes from A to B, you need to reverse the velocity of A. When we try to find a difference, we always subtract the last one from the first one.



When considering the triangle of PQR, $-v_i + v_f = \Delta v$ ($\overrightarrow{RQ} + \overrightarrow{QP} = \overrightarrow{RP}$)

$\Delta v = v_f - v_i$. As v_f and v_i are equal in magnitude, the magnitude of Δv is $\sqrt{2} v$. The direction is towards RP direction.

However, the acceleration of the object is directed towards the direction of Δv as $a = \Delta v / \Delta t$; We know that the acceleration of an object which moves in a circular locus at a uniform speed is directed towards its centre. Therefore, Δv should be directed towards a direction like this.

As the acceleration is directed towards the centre, the net force also should be directed towards the centre. If we do not have an affection or a love towards a person in the middle, then we cannot go around the person in the middle.

09. A weightlifter lifts a weight vertically up (positive direction) with his hands.

The signs of the work done by

- (a) his hands on the weight, (b) gravity on the weight, and
(c) the weight on his hands, respectively, are

	(a)	(b)	(c)
(1)	+	+	+
(2)	+	-	+
(3)	+	-	-
(4)	-	+	-
(5)	-	-	+

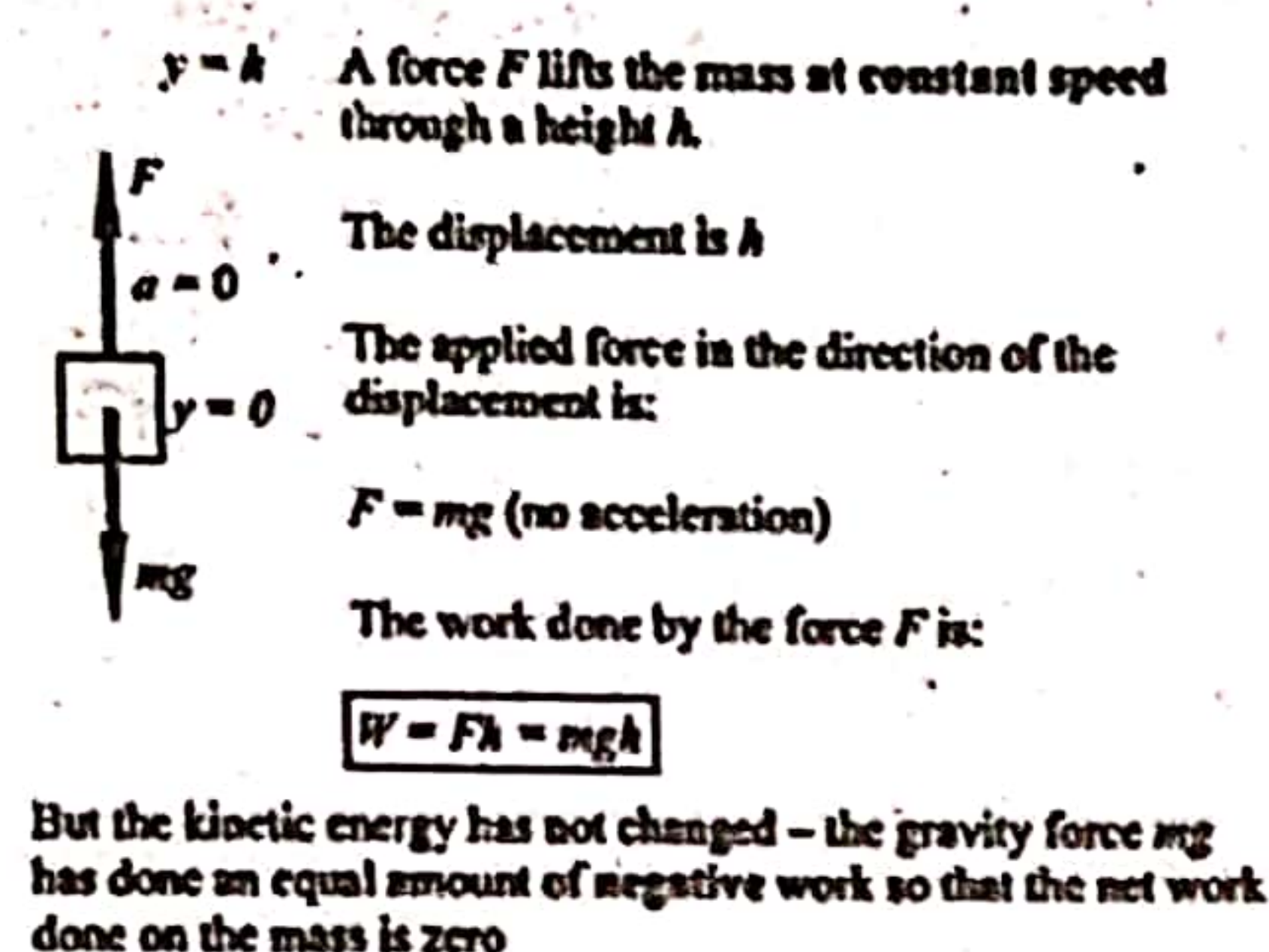
Friction

02

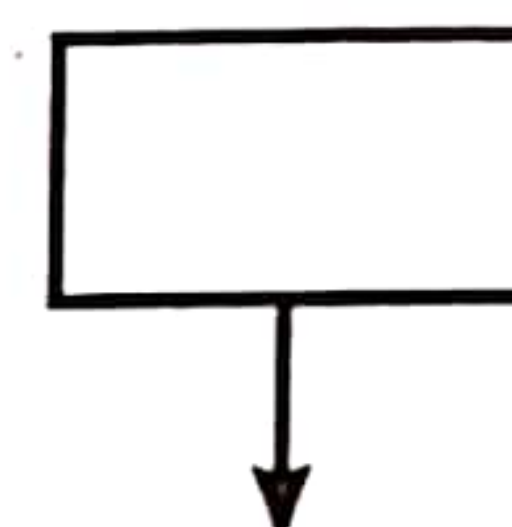
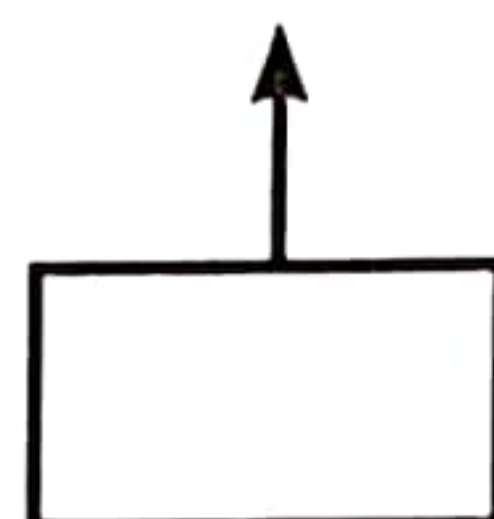
Think of a person who is lifting an object. To lift the weight, he needs to apply an equal and opposite force (F) to mg upwards. If we consider the displacement upwards as positive, then the work done from his hands will be positive (+). $\uparrow F \uparrow$ displacement

The gravitational force is acting downwards. The displacement occurs upwards. Therefore, the work done from the gravity is negative (-). $\uparrow mg \uparrow$ displacement. However, the man is doing work against the gravitational forces. The force that acts on the object by the hands (F) is acting upwards. According to Newton's third law, the force on the hands by the object is downwards $\downarrow F$. Therefore, the work from that force is negative (-).

Work done in lifting an object



F F (to object by hand)



F (to hand by object)

10. Consider the following statements regarding a three-level LASER system having energies E_1 , E_2 , and E_3 ($E_1 < E_2 < E_3$) as shown in the figure.

(A) The LASER action occurs between energy levels 2 and 1.

(B) The frequency of pumping radiation is

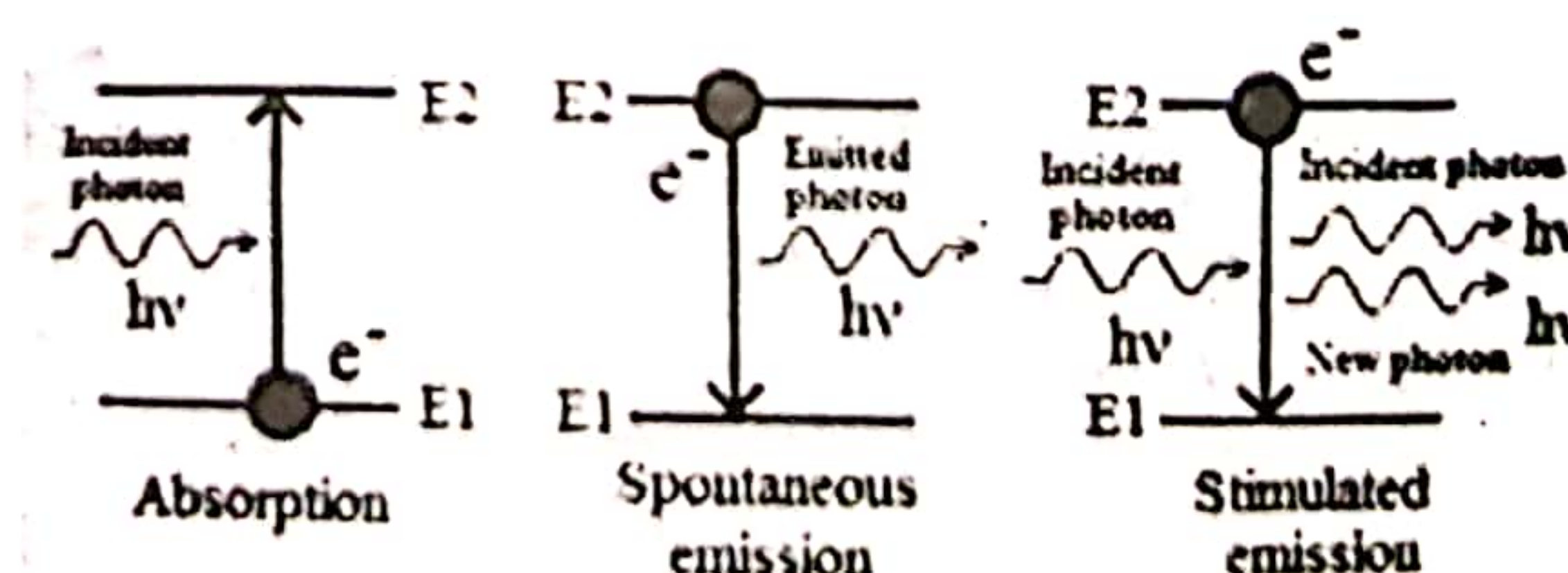
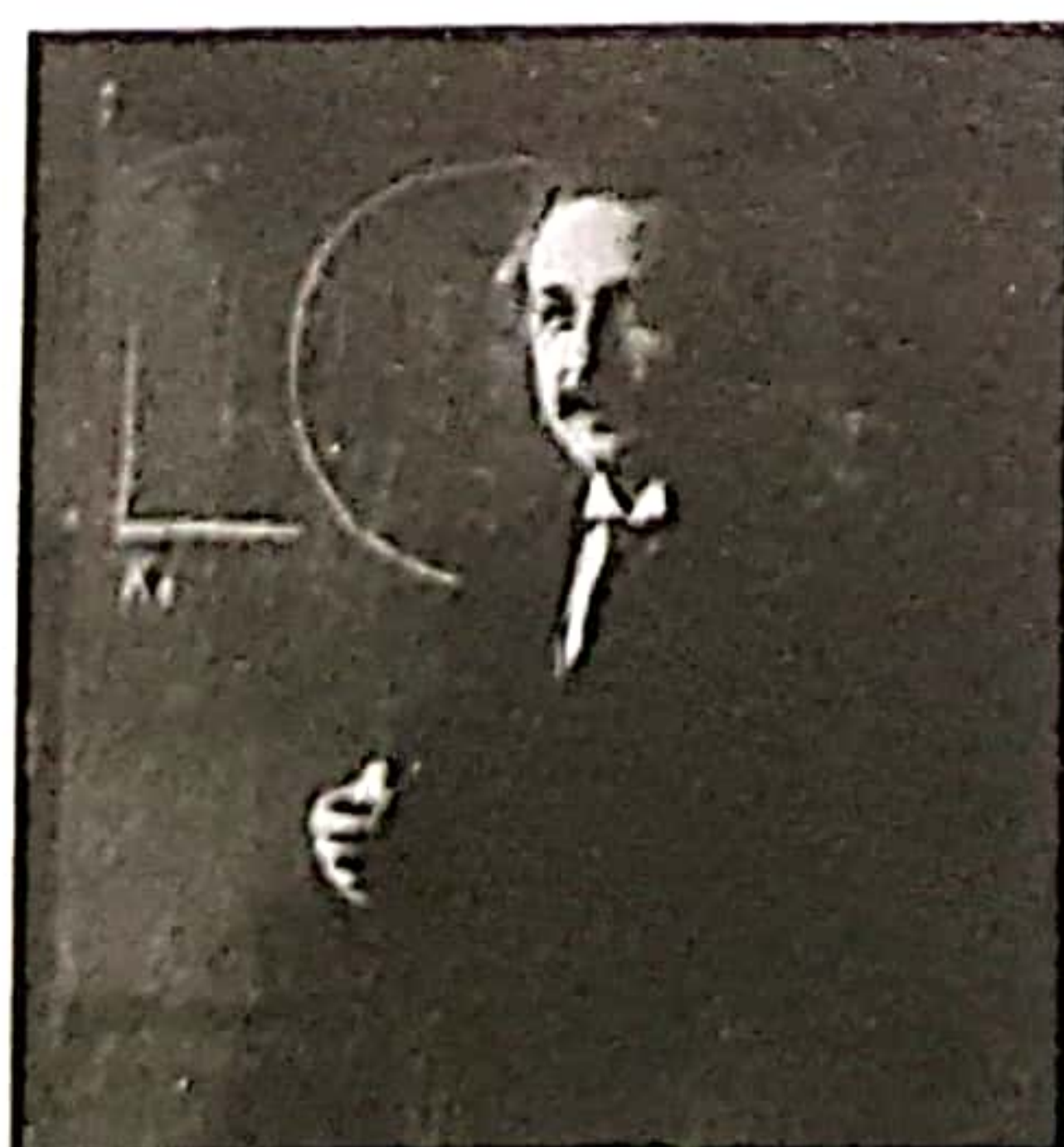
(C) Level 3 is known as the metastable energy level. Which of the above statements is/are correct?

- (1) Only A (2) Only B (3) Only C
(4) Only A and C (5) Only B and C

03

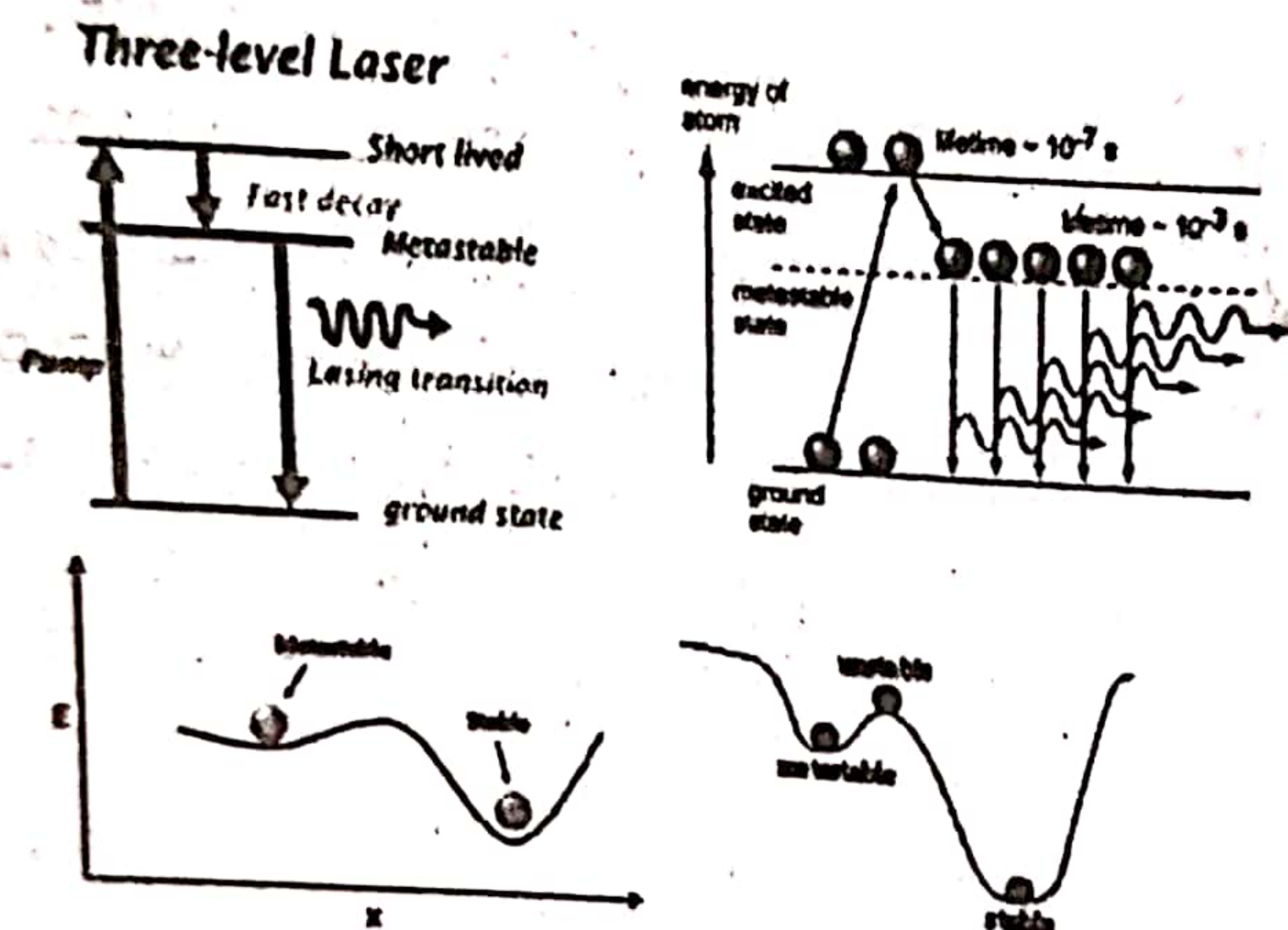
Electromagnetic Waves

The meaning of laser is Light Amplification by Stimulated Emission and Radiation. The figure has shown the absorption, spontaneous emission and stimulated emission processes. An electron that is in a lower level of an atom can be sent to a higher energy level by a photon with a certain frequency (with the energy difference of two levels). This process is known as the absorption. When the electrons are transferred to higher energy levels, the atoms get unstable. When there is higher energy, it tries to release the energy. This is common to nature. Such excited electrons are normally fallen to the lower level in a very short time (10^{-7} s- 10^{-9} s). This random process is known as the spontaneous emission.



In 1917, Einstein mentioned that there can be an occurrence of another process here. That is when the

electron is at the upper energy level, the photon which has the energy equal to the energy difference of two levels can excite the electron to come to the lower energy level. This is known as the stimulated emission. The incoming photon excites the electron to drop into a lower energy level. What happens here is that without an energy difference in the excited photon, the photon from the stimulated emission goes with the initially arrived photon together. Even humans like us can be excited for good or bad. The exciting person is like a catalyst. His/her energy is not wasted. In the stimulated emission, we get both the person who excites as well as the product of excitation.



Therefore, it is a kind of amplification. The emitted photons excite many other electrons in the atoms and get an army of (monochromatic) same energy.

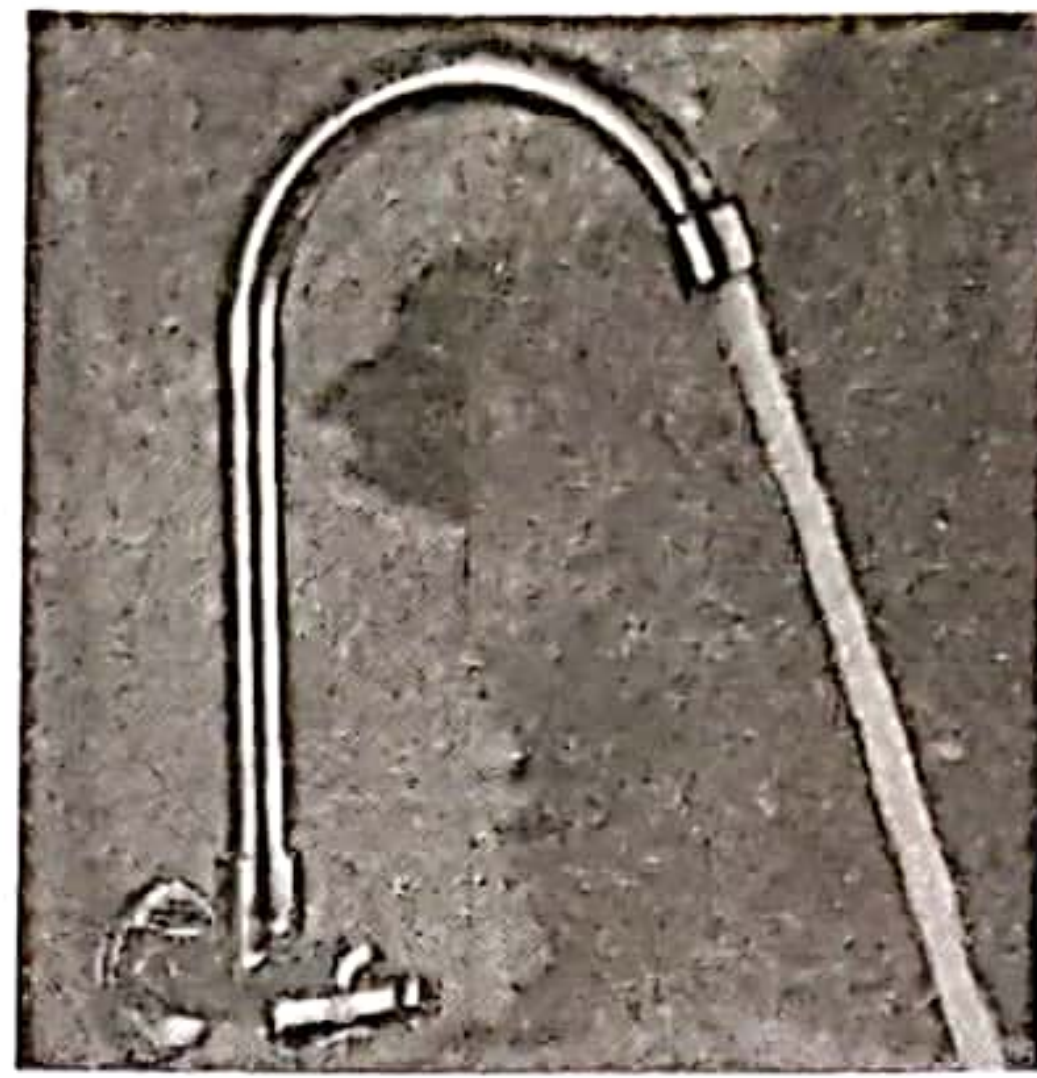
Now we will consider a three-level laser. First, the electron is kept at the third level from the ground level. This is known as pumping. This word itself has the real meaning of it. The selected electrons that go to third level should be transferred quickly to the second level. That means from spontaneous emission, the electrons should be transferred (decay) from the third level to the second level in a very short period (s). ($\sim 10^{-7} \text{ s}$).

A metastable energy level should be taken as the second level. A metastable energy level means a level where the electrons stay for a while without falling quickly. As mentioned early, the life time of electrons which undergo spontaneous emission is about in the range 10^{-7} s - 10^{-9} s . But the electrons that live in metastable level has a life time of nearly 10^{-3} s . You may wonder that even 10^{-3} s is a small time. But 10^{-3} s is bigger by 10^4 times when compared to 10^{-7} s . For the process of laser, compulsorily there should be a metastable energy level. The reason is that for the occurrence of stimulated emission, the electrons should be kept at an energy level like this for some time. If there are nobody to excite, then who are going to be excited? The persons who quickly fall cannot be excited.

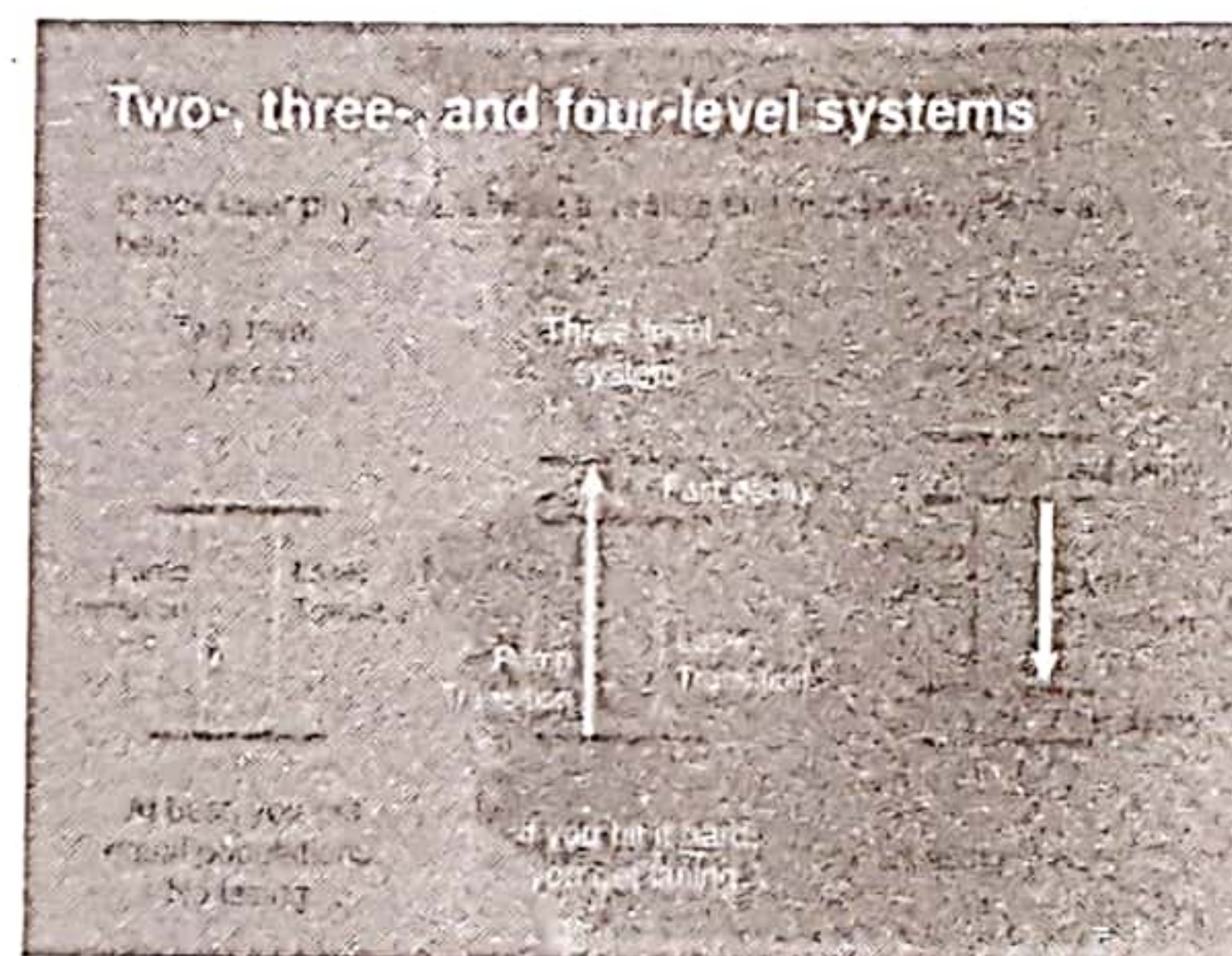
A mechanical state that represents a metastable energy level has been shown in the previous figures. The stable state ball is in the ground. The metastable level ball is at a shallow hole. It is neither too deep or shallow. Due to corona, do our children are in the metastable levels? But there is a hole nearby to fall once it is filled. When waves like UV are fallen into fluorescent materials, they emit light. But when the incident energy is stopped, the fluorescence stops. But even when energy like UV is incident on phosphorescent materials and the source get deactivated, the emitted light will not spontaneously stop but will gradually reduce. The reason is that, there are metastable energy levels to phosphorescent materials. Zinc sulfide, Strontium aluminate are examples for phosphorescent materials.

Some gets angry quickly. Some cools down the anger quickly. Some keeps the anger in themselves for a longer time.

Now when the pumped electrons are fallen into third level, they will quickly fall into second level and fill it. The third level is like a bucket which has a big hole at the bottom. When water is filled into the bucket, it falls quickly into the bucket at the second level. That bucket has a small hole at the bottom. Therefore, water falls in drop wise to the first level. It will not fall quickly.



There is another special fact to the laser process effectively. That is, compared to the lower energy level (first energy level) there should be more electrons in the metastable level. The lower-level population should be less by allowing the upper-level electrons to fall. This is called as population inversion. Normally, electrons are situated in the lower energy levels of the atom. But to achieve the laser process, the metastable level should have more population compared to the lower energy level. This is not the normal situation. It is an inverse situation. That is why we called this phenomenon as the population inversion.



The figure shows the systems of level two, three and four. The laser process cannot be done only from two levels. From that, the previously mentioned population inversion cannot be achieved. It is like the water fallen from a waterfall to a lower level is put back to into the upper level. Putting and taking or in other words taking what is put. Therefore, to produce laser, you need at least three levels as shown in the second figure. From 1 to 3, the pumping is done. From 3, the electrons fall to 2 quickly and fill 2. The laser process (stimulate emission) is happened from 2 to 1. From this you will get the answer. Only statement (A) is correct. The frequency of the pumped radiation is $(E_3 - E_1)/h$ not $(E_3 - E_2)/h$. The metastable level is 2 not 3. A system with four levels can do the laser process conveniently and more efficiently. Why is that? From 1 pumping is done to 4. From 4, the electrons fall quickly to 3. 3 is the metastable level. It protects the received ones. It will contribute to the stimulated emission and falls 2nd level. Now whoever came to 2 quickly falls to 1. Therefore, the population inversion condition is very well satisfied in level 3 and 2 (the laser process occurs from 3 to 2). Level 4 is a bucket with a bigger hole at the bottom. Level three has a smaller hole and again level 2 is like a bucket with a bigger hole. When water is filled, it pushes the water to level 1. There is no hole in the bucket of level 1. There is no usage even if there is a hole as it is the ground state. There is nobody to fall below it. What can happen is the lifting to the level 4. This energy is given by an external source.

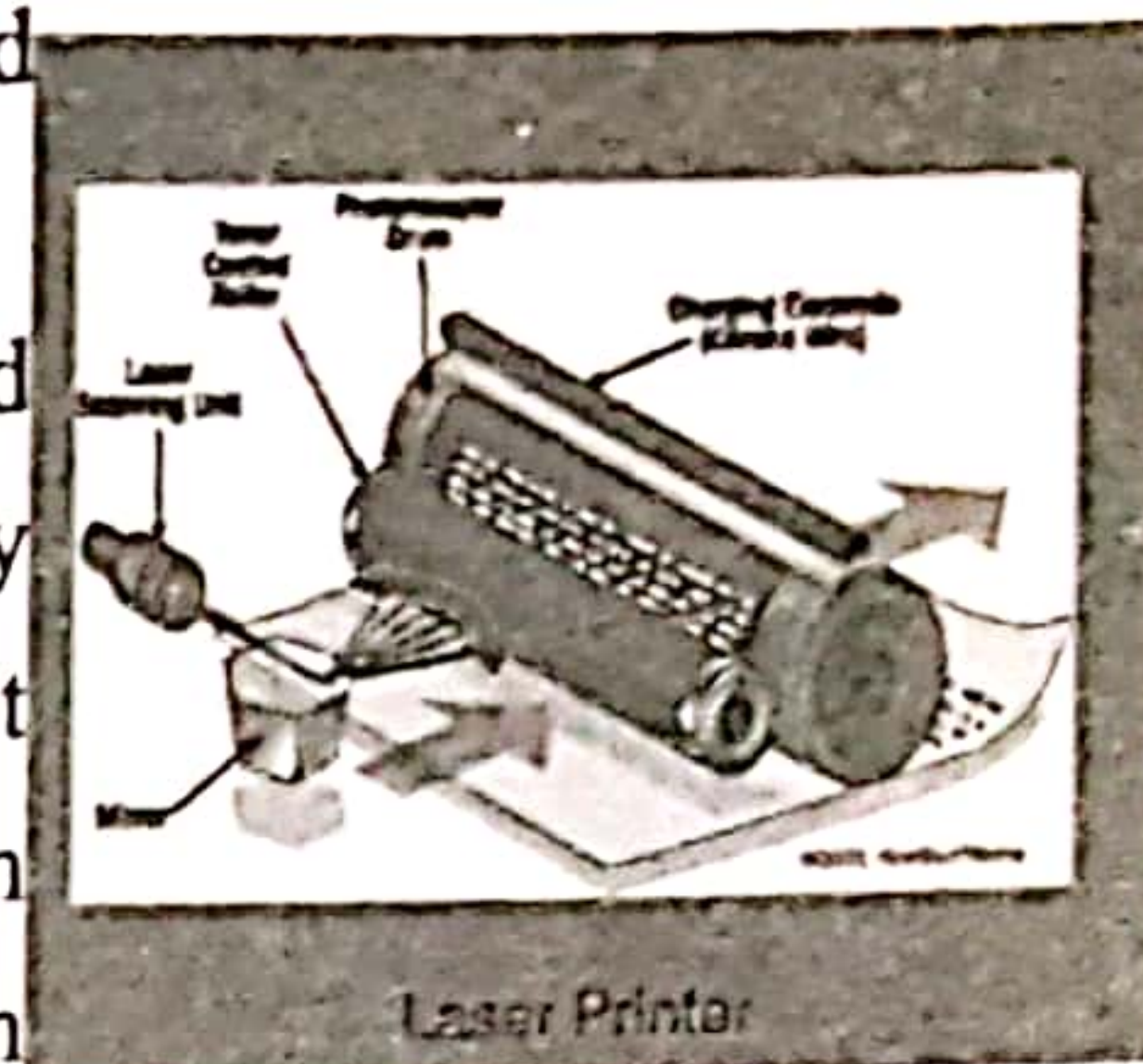
There is another fact that you should know. The energy level difference between 4 and 3 as well as 2 and 1 should be smaller. If light is emitted from the lase transfers from 3 to2, then from other transfers light should not be emitted. If so, then the process is not effective. It does not matter if heat is generated in those transfers.

That is why level 4 and 3 as well as 2 and 1 should be kept nearby.

When creating a laser material, all these facts should be satisfied. Sometimes, you may not get all the properties from one material. Some materials are mixed when necessary (He-Ne Laser). Study of them is bit complex and beyond the A/L syllabus.

Usage of laser is common in every field. The laser technology is being used in almost every field from a laser pointer to reading bar code of a goods, scanning the information in a CD, laser printing machines, 3D printing machines, from cutting of metal plates of machines very delicately and accurately to medical procedures in laser eye clinics/ eye surgery.

If there are unwanted hair in our daughters' bodies, then they can be removed or permanently stop growing with the usage of laser technology. Einstein may have not thought in his dreams about hair removal in ladies with laser. The root of the hair can be closed and seal from laser here. As mentioned earlier, Einstein first introduced the stimulated emission concept and told that it can be done in 1917. Charles H. Townes was the person who studied many theoretical things and develop laser further.



In 1964, he was awarded the Nobel Prize for Physics with some other scientists. When he was resting in a park, he noticed how small water droplets that had fallen to a leaf combined together and flow along the main stem. He was convinced that likewise, photons can be combined together to make a single laser beam. The first laser was made by Theodore Maiman on 16th May 1960.

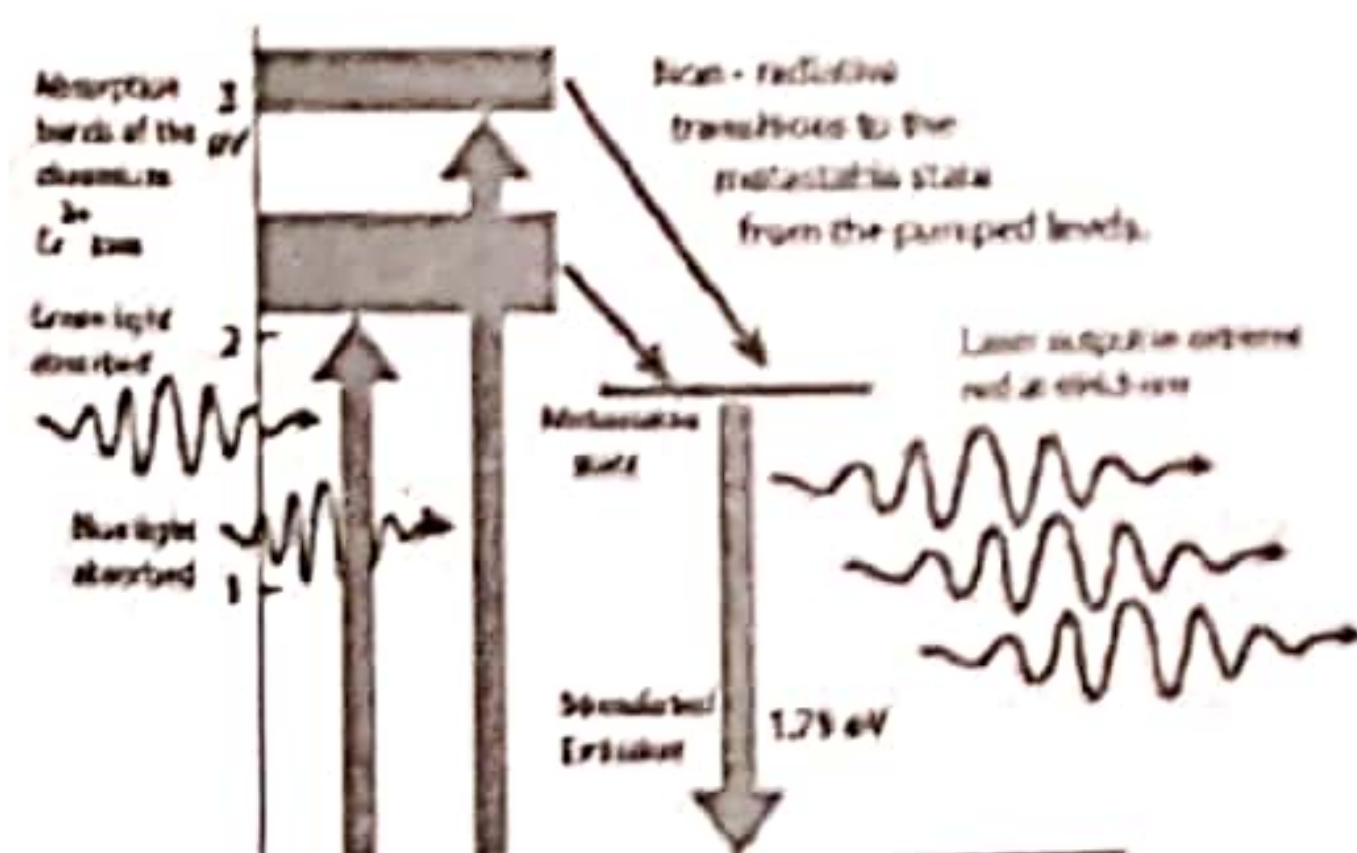


His laser medium was Ruby crystal material. Ruby is sapphire crystals that we know (It has crystalized aluminium oxide and a small 0.05% percentage of chromium).

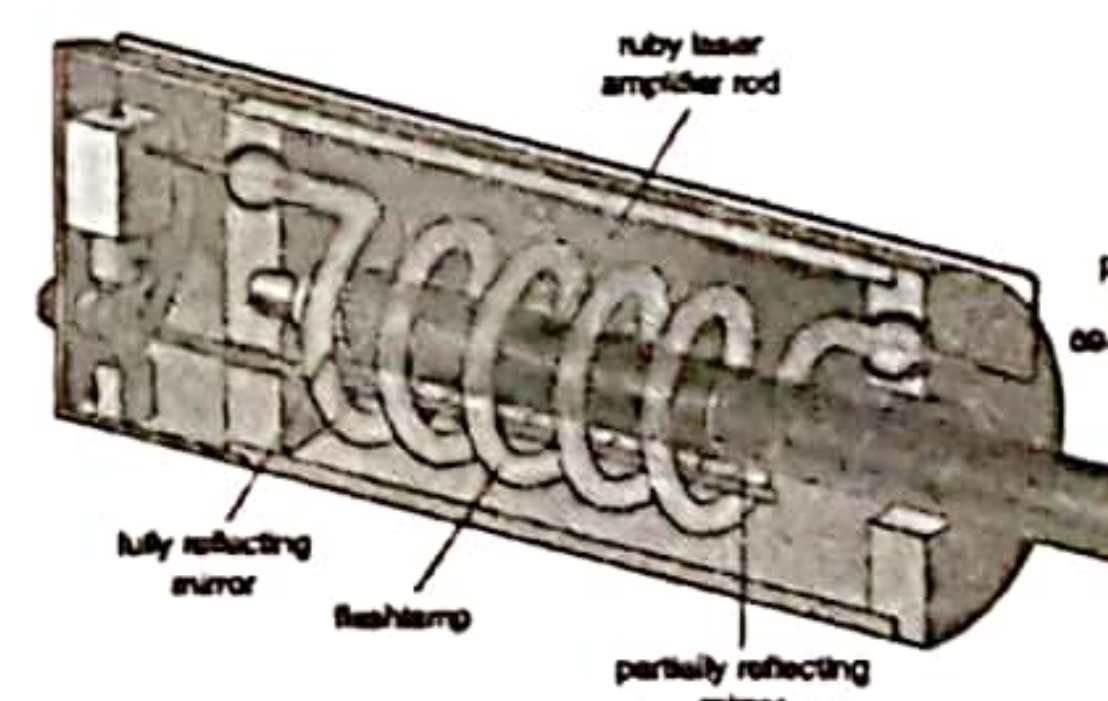


WORKING OF RUBY LASER

* Ruby laser is based on three energy levels. The upper energy level E₃ is short-lived, E₁ is ground state, E₂ is metastable state with lifetime of 0.603 sec.



This emits laser light in red colour. To pump electrons, a flash light is used. The three energy levels of the crystal have been shown in the figure. Ruby laser is considered as a three-level laser. But the metastable level is fed from two higher energy levels. Even the flash light emits blue light from one pumping and green light from the other pumping.



11. Consider the following statements made regarding the velocity of sound in earth atmosphere.

- (A) It does not change with altitude at constant temperature.
- (B) It always increases with decreasing pressure.
- (C) It decreases with increasing altitude as a result of decreasing temperature. Which of the above statements is/are correct?

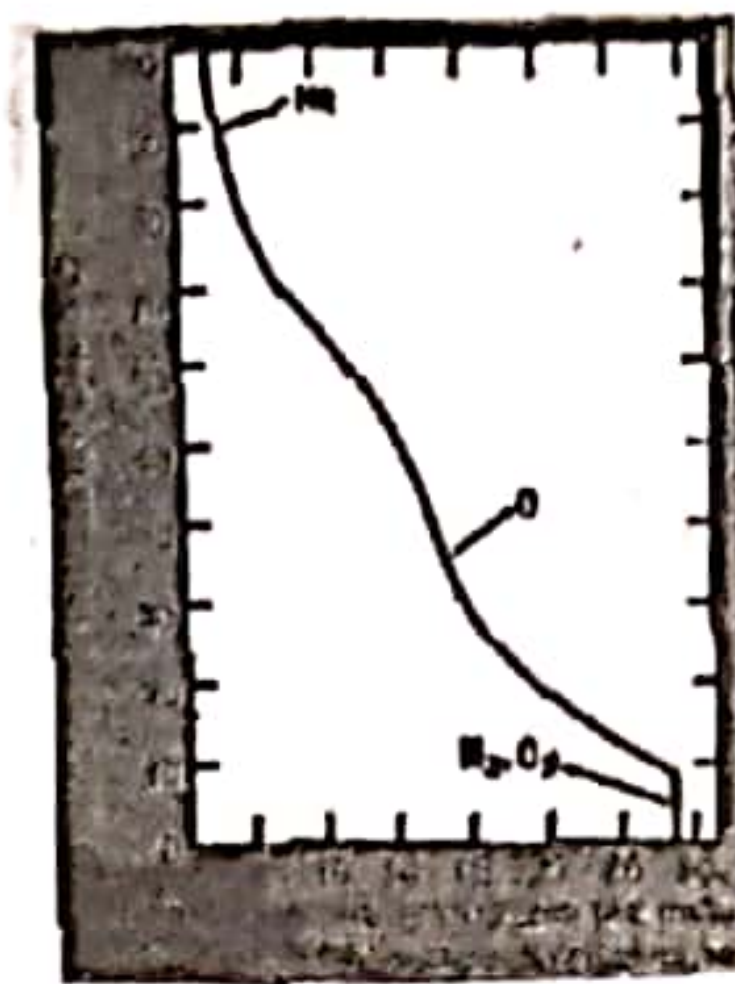
- (1) Only A
- (2) Only B
- (3) Only C
- (4) Only A and C
- (5) All A, B, and C

Velocity of Sound

03

As the sound speed of the air given by the equation $v = \sqrt{\frac{\gamma RT}{M}}$, for a given gas $v \propto \sqrt{T}$. This has been checked many times. The Sinhalese word for altitude can be an unfamiliar word for children (උන්නතාංශය). This word may be unfamiliar to children as this word has not been used in previous papers. This can be given as the height measured from the ground.

If the temperature is considered constant as we go upwards, then the sound of speed cannot be changed. But some children questioned that when going upwards from ground cannot v be changed as the net value of M can be changed due to the change the composition of the atmosphere? When finding more about this matter, it was noted that the scientists have found the composition of the atmosphere is unchanged till 10 km from the ground. The composition of the atmosphere (M) with the altitude variation is shown by this figure.

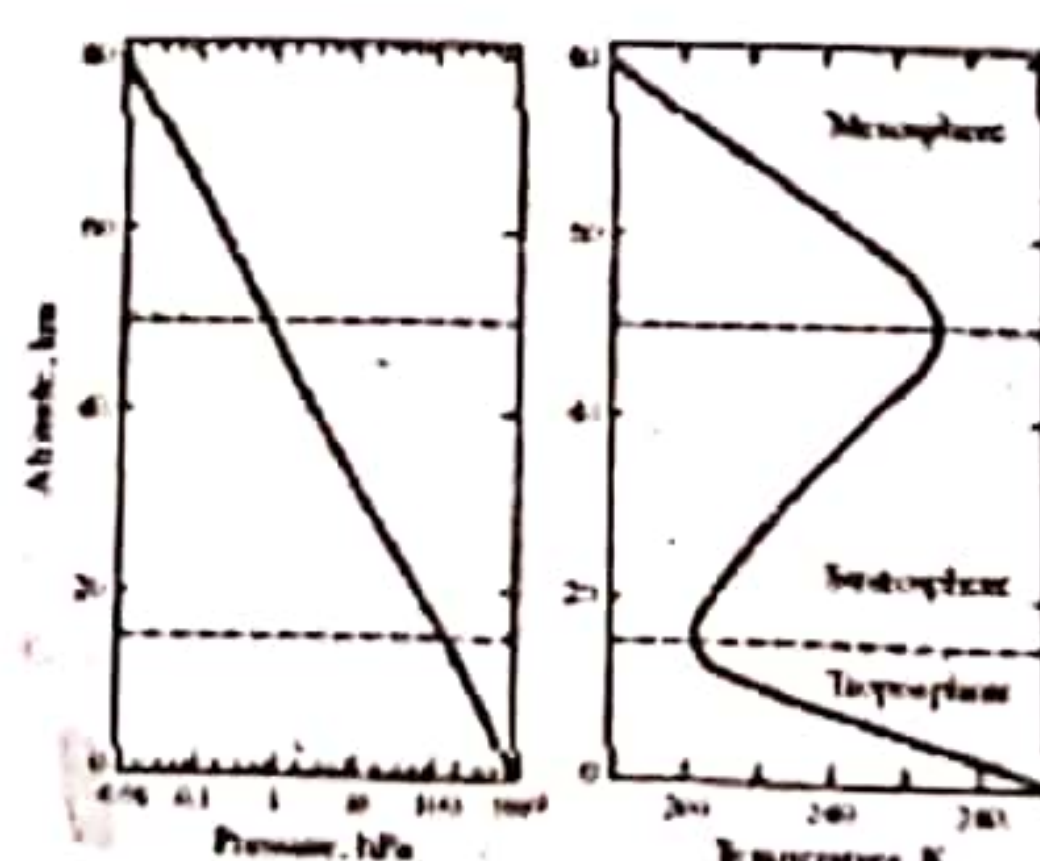


Until 10 km height, there are mainly molecular Nitrogen (N_2) and molecular Oxygen (O_2). Once it goes beyond 10 km, the percentages of N_2 and O_2 get reduced and atomic oxygen (O) and triatomic oxygen (O_3) get prominent. Therefore, as 10 km is a considerable height, I feel that it is fair to consider that γ and M values remain unchanged. That means if temperature is constant, the statement of sound speed does not change with the altitude can be considered as true. But it is not true when it passes 10 km. However, once we go upward with the altitude, clearly the temperature will not remain constant. The constant temperature is an assumption only. (The figure that is taken from the Internet has mentioned the altitude axis as 100 km, 200 km, 300 km.. I have shown the same figure in the video show too. But it should be corrected as 10 km, 20 km, 30 km..)

Some intelligent children think a lot of these issues. It is not a bad practice. It is fair if someone argues as there is no mention about the altitude limit from the ground and this statement is wrong. I feel it is better to consider about this issue.

However, the speed of sound is not varying with the pressure. This has been given many times. According to $v = \sqrt{\frac{\gamma P}{\rho}}$, when P is changing ρ is also changing. However, $v = \sqrt{\frac{\gamma RT}{M}}$ equation is obtained by the substitution of P/ρ in the equation of $v = \sqrt{\frac{\gamma P}{\rho}}$,

In the third statement, it has been mentioned that the temperature decreases with the altitude. If so, then the speed of sound should be reduced. In the first statement, the temperature is constant. In the third statement, the temperature is reduced. First and third statements have been considered as correct. The atmospheric pressure and temperature variations with the altitude has been shown in these graphs.



I feel that the assumption of temperature remains constant is unfair. But can we assume anything in Physics?

In troposphere, the temperature reduces when it is away from the ground. The reason is that the gases in this region (N_2 , O_2 , CO_2) absorb the long wavelengths of radiation from earth heating instead of absorbing solar radiation very much. The other fact is that when the air goes upwards due to convection, the volume is increased as the pressure is decreased. This is considered as an adiabatic (quickly happening) expansion.

In an adiabatic expansion, [$\Delta Q = 0$; $\Delta U = \Delta W$; ΔW is positive (as the volume is increased)] the gas is getting cooled.

In stratosphere, the temperature is increased with the altitude. The reason for this is the generated heat due to the absorption by the ozone layer from the incoming UV rays of the sun.

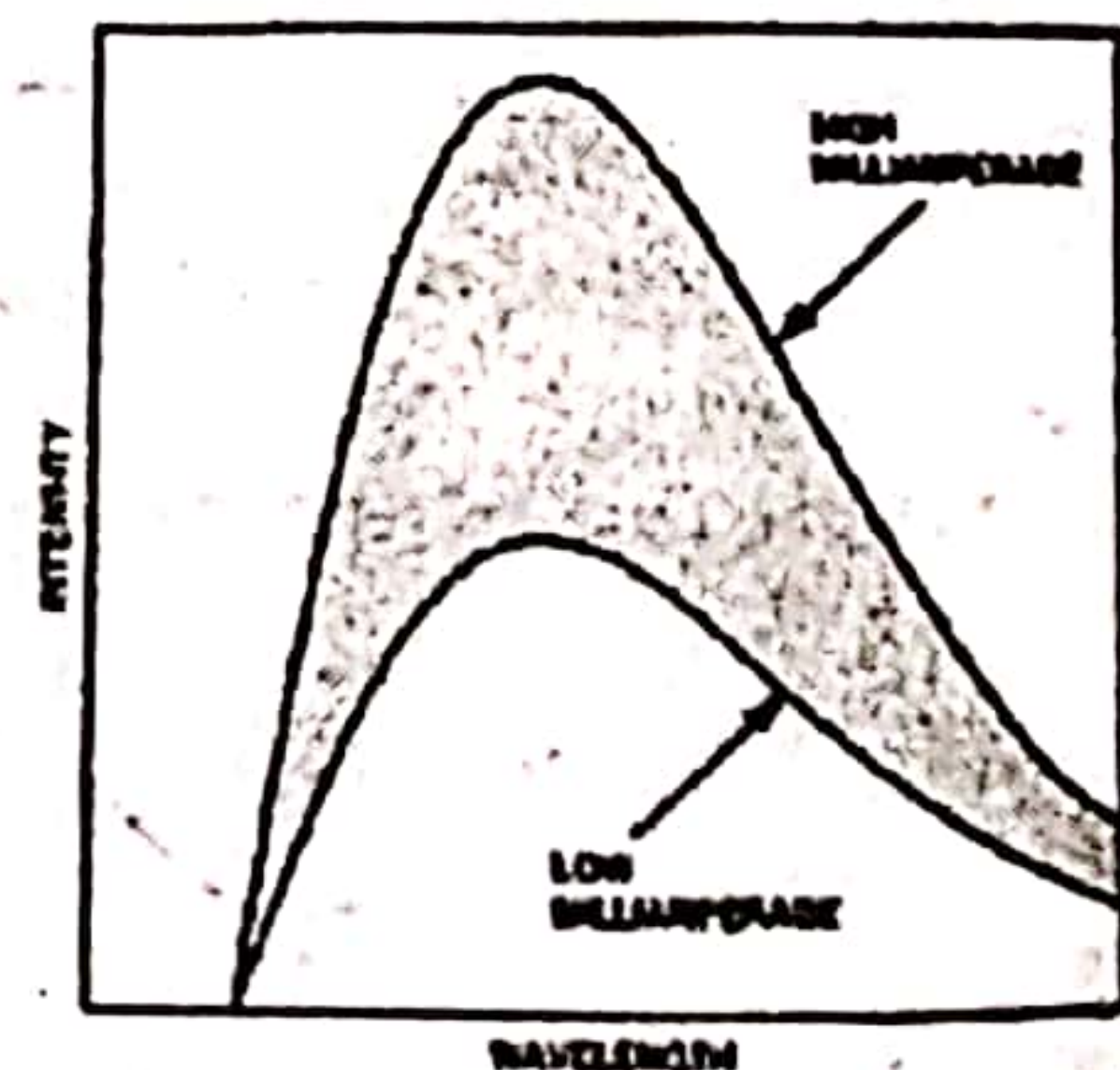
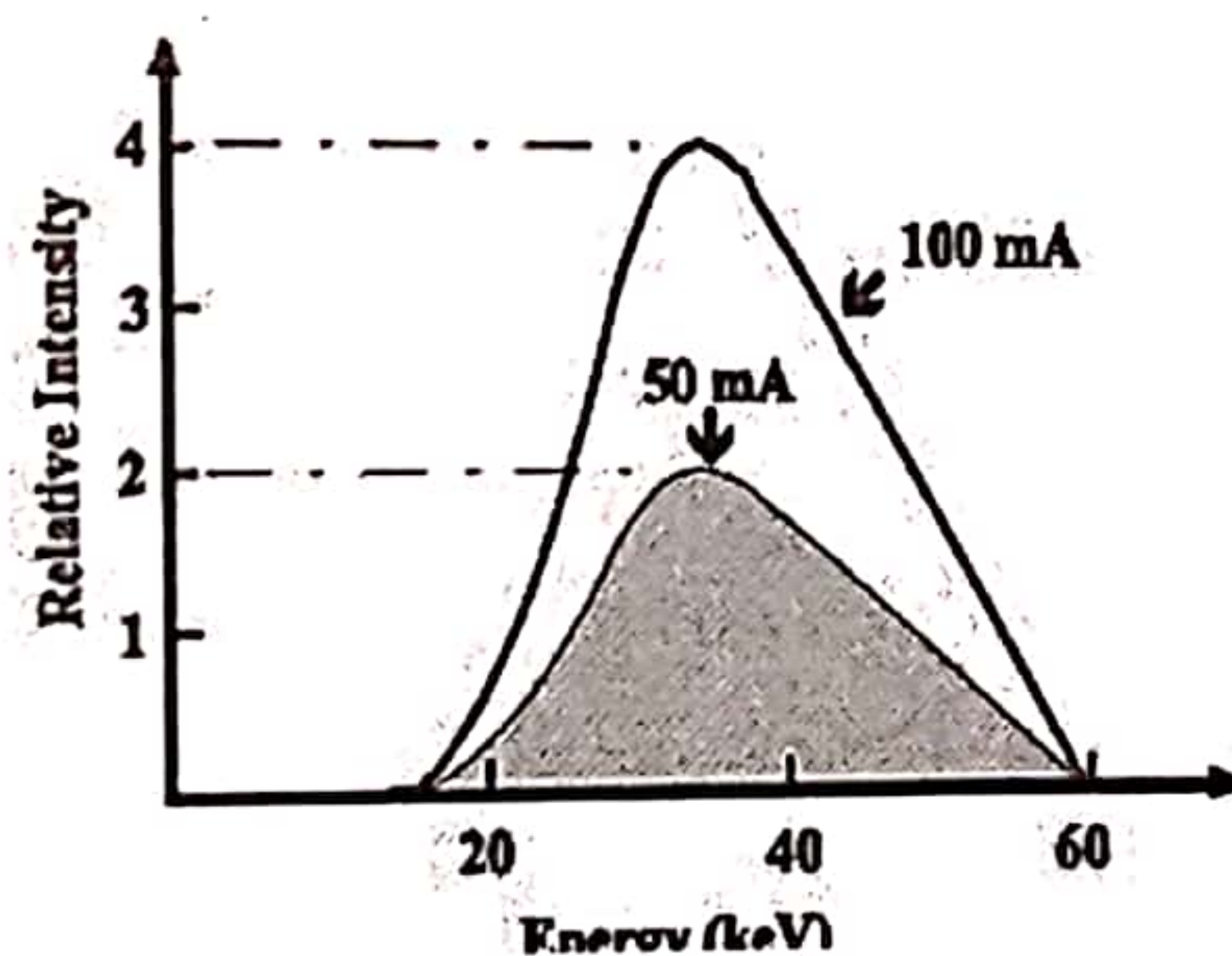
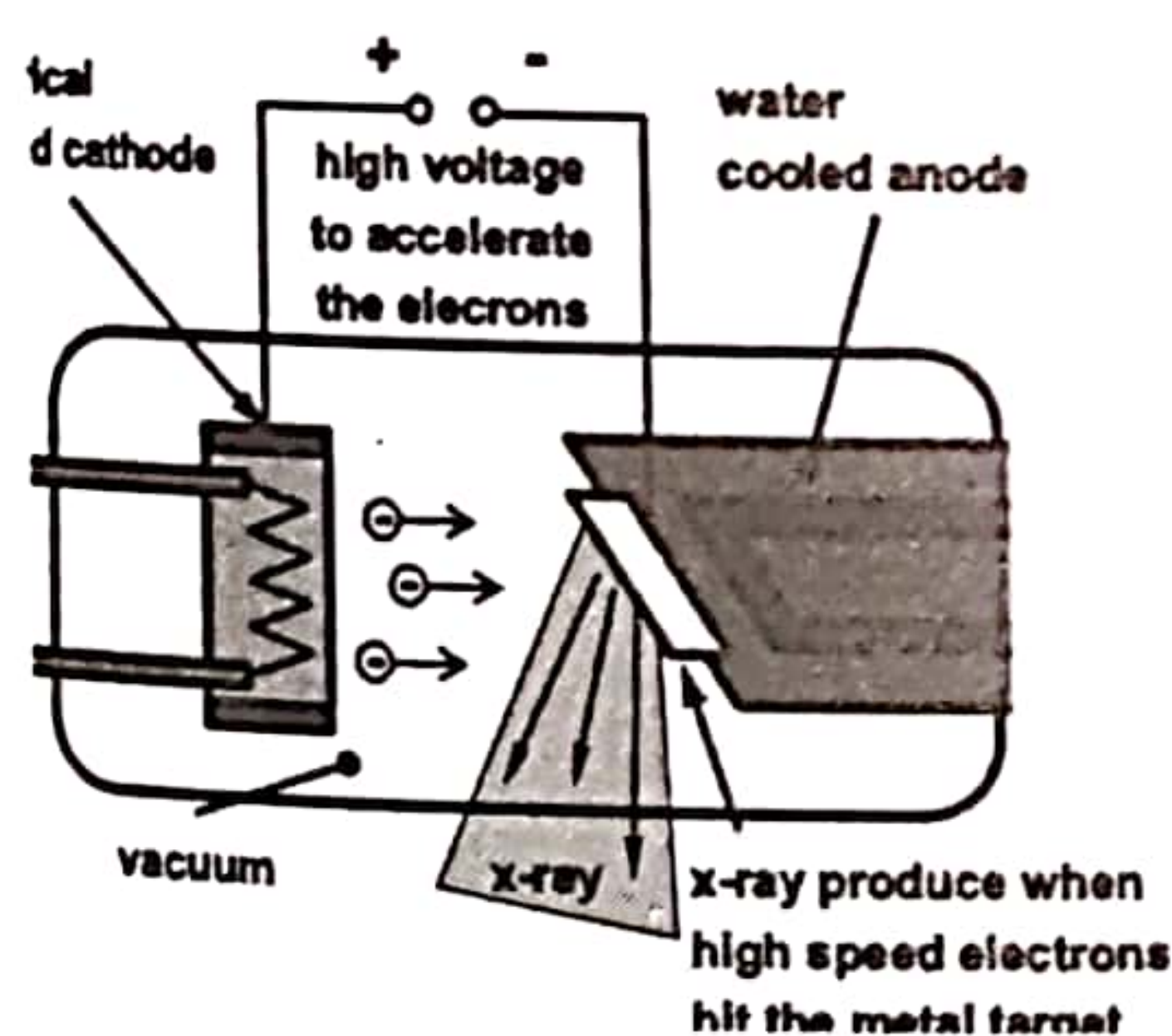
In mesosphere, there is not much air in this region. The gas density is reduced when it goes upwards. Therefore, the temperature also gradually reduced.

12. Which of the following statements regarding the X-ray production in common applications is incorrect?
- (1) Two circuits are used in the X-ray production system.
 - (2) Anode could be damaged due to the bombardment of electrons.
 - (3) Low voltage is sufficient to heat the cathode.
 - (4) Energy of the X-rays emitted depends on the current through the filament.
 - (5) X-ray tube must be evacuated to avoid the energy loss of electrons.

Particles and Waves

11

I have written a detailed explanation about X-ray tubes and the production of X-rays, in the review of 2017. There should not be air in the X-ray tube. If there is air, then the electrons will scatter here and there while losing the energy in them. There should be two voltage supplies to heat the filament of the tube (cathode) as well as to keep a potential difference between the anode and the cathode. That means you can consider that there are two circuits. To heat the filament, a considerable amount of current should flow in it. The filament consists of a coiled wire (normally tungsten) and it gets incandescent by increasing its temperature. The filament emits electrons by thermionic emission. A low voltage is enough to heat the filament. All you need is to allow a high current (3-7A) flow across the filament. This can be obtained by using a step-down transformer.



There should be a high voltage (in kV range) in between the cathode and the anode. Otherwise, the kinetic energy of the electrons cannot be increased into keV range. The electrons are hit to the targeted material. The anode can be considered as the targeted material. Most of the time, the targeted material (tungsten) is embedded on the surface of a copper rod. Actually, the damage can be happened to the target. The target can be damaged due to the higher heat energy due to electron collision.

When the filament current is increased, more electrons are emitted from the filament. But the kinetic energy of the hitting electrons on the target does not change. The kinetic energy of the electrons can be increased by increasing the voltage in between the cathode and the anode. But when more electrons are hit on the target, the amount emitted X-ray photons are increased. But the emitted X-ray photon energy is unchanged.

Actually, the intensity of X-rays is dependent upon the current that flows in the filament. The penetrating power of X-rays (the energy of a photon) is dependent upon the voltage in between the anode and cathode that supports the acceleration. Therefore, the statement of 'the energy of an emitted photon is dependent

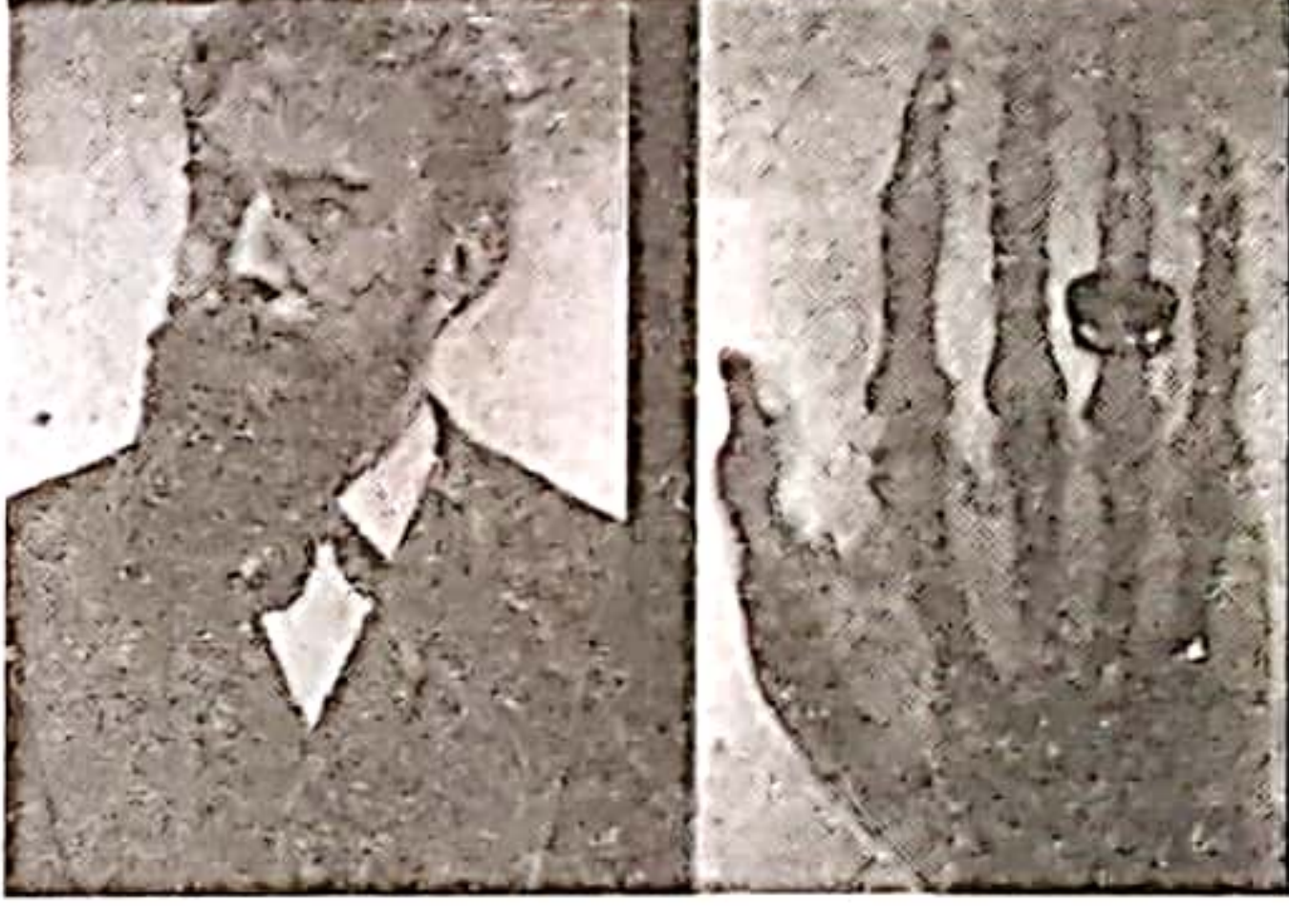
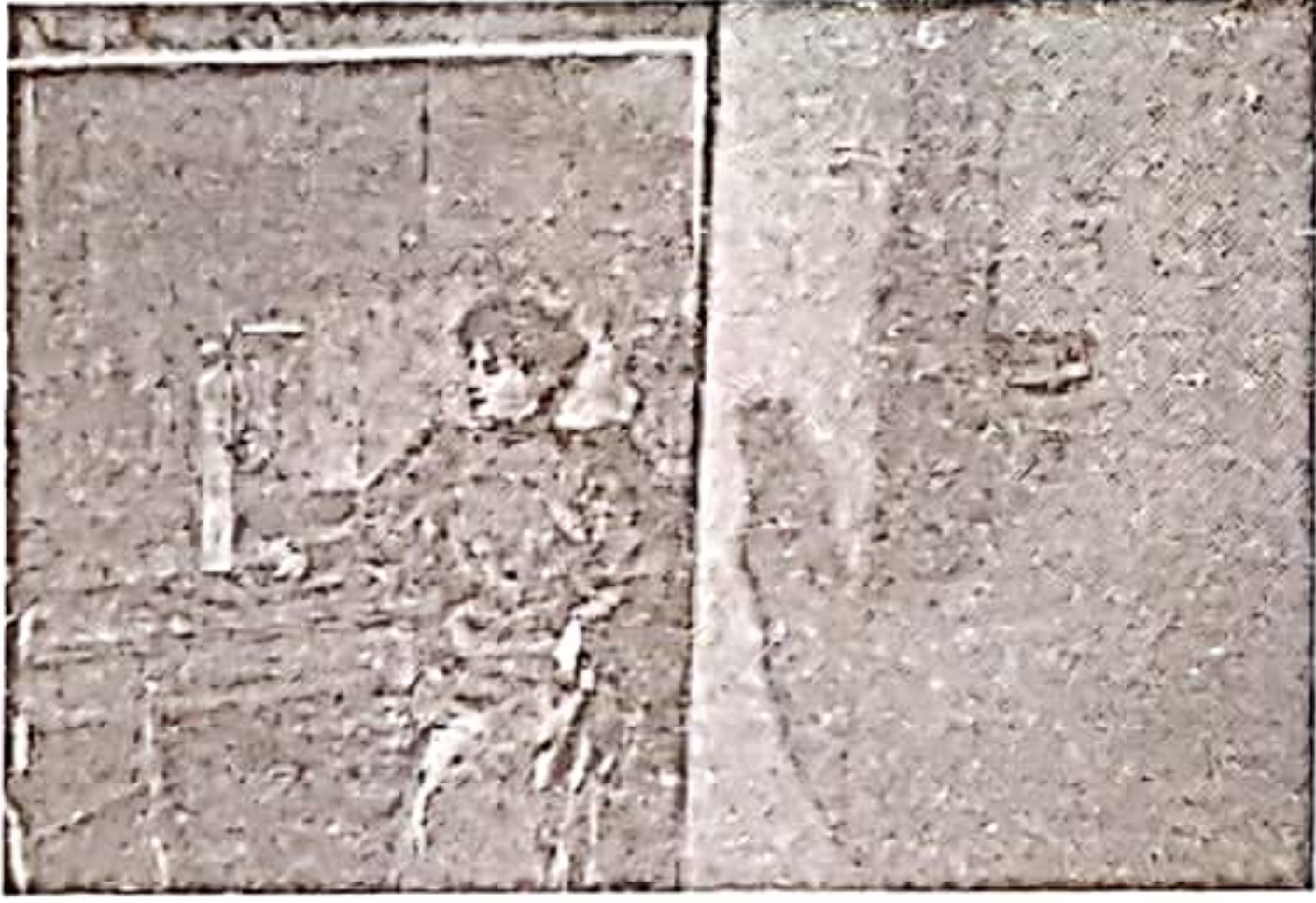
upon the filament current' is clearly false. When the filament current is increased, even though the energy of X-ray photon remains unchanged, the number of X-ray photons get increased. That means the intensity of the X-ray beam is increased.

If you increase the filament current by keeping the acceleration voltage constant, then more and more X-ray photons are emitted. If the acceleration voltage is increased by keeping the filament current constant, then the energy of the emitted X-ray photons is increased.

When both filament current and the acceleration voltage are increased, the number of photons that are emitted as well as the energy of each one is increased. It is better to write the energy of X-ray photons instead of the energy of X-rays. Look at the following figures. The left figure shows Rontgen and his wife Anna.

When Rontgen was working with the cathode ray tubes, he observed that unknown rays are emitted with a penetrative power from the place where the electrons are hit. He noticed that a crystal which was on the table near the tube, suddenly showed fluorescent glow. He was in the laboratory for 6 weeks and had studied these unknown rays. Sometimes his wife might have come in searching for him. The first X-ray of the world has shown from the right side. Sometimes the wife voluntarily might have participated in it. Otherwise, Rontgen might have begged. However, men do not allow wives to poke into unknown things. When she saw the shape of her bones, she had said that "I have seen my death" to a newspaper in 22nd December 1895.

Normally, we see the skeleton after the death. Rontgen won the first Physics Nobel Prize in 1901 for this new discovery. The foundation for Nobel Prizes was established by Alfred Nobel. That was from the money he obtained by selling



dynamite.

The study and use of ionizing radiation in medicine started with three important discoveries

- ✓ X rays by Wilhelm Roentgen in 1895.
- ✓ Natural radioactivity by Henri Becquerel in 1896.
- ✓ Radium-226 by Pierre and Marie Curie in 1898.



13. Consider the following statements regarding the dew point of air having water vapour in a closed container.
- (A) At dew point, unsaturated water vapour becomes saturated water vapour.
 - (B) If the temperature is reduced below the dew point, some of the vapour will condense.
 - (C) At dew point, if the volume of the container is reduced, the absolute humidity of the air will decrease.
- Which of the above statements is/are correct?
- (1) Only A
 - (2) Only B
 - (3) Only A and B
 - (4) Only A and C
 - (5) All A, B, and C

04 Hygrometry

When the temperature of the air is being reduced, at a certain temperature, the space can be saturated with the water vapour in air. That temperature is known as the dew point temperature. Therefore, when the temperature is reduced in a closed container, the unsaturated water vapour turns into saturated water vapour at the dew point. If the temperature is reduced than the dew point temperature, part of the water vapour tends to condense. This is also correct.

When the air is being saturated with water vapour and the volume of the container is reduced, then certain water vapour amount will be condensed. Therefore, the mass of water vapour in the container gets reduced. But volume also gets reduced. Therefore, absolute humidity (the water vapour mass in a unit volume) is not changed.

In a closed space, if we reduce the temperature without changing the volume, then the absolute humidity of the space gets reduced. As part of water vapour is being condensed, the mass of water vapour in the space is reduced. As the volume is unchanged, the (remaining mass/ volume) ratio is being reduced.

These facts (without the change of volume) have been checked many times in the past papers (2000, 2008). If there are unsaturated water vapour in a closed room, then when the temperature is reduced the relative humidity is increased. But the absolute humidity is at a constant value. When the temperature is equal to the dew point temperature, then the relative humidity gets 100%. If we reduce the temperature lower than the dew point temperature, then the relative humidity remains in 100%. But as the water vapour gradually get condensed, the absolute humidity gets reduced.

Another fact that you can see is that when the space of the container is saturated with water vapour, the saturated water vapour pressure is not changed depending upon the volume. Normally, when the volume is reduced the pressure should increase (for a constant mass). But when the volume is reduced, the water vapour is turned into liquid state from the vapour state. Therefore, the increment of pressure is cancelled off. According to $PV = nRT$, if T is constant and while reducing V if P is constant, then n/V ratio should be a constant. V is reduced and simultaneously n (m) is also reduced. Then the absolute humidity $[n/V(m/V)]$ is constant.

14. When the tension of a wire is slowly increased from T_1 to T_2 within the proportional limit, its length changes from l_1 to l_2 . The energy stored in the wire during this process is

- (1) $(T_2 + T_1)(l_2 - l_1)$ (2) $\frac{1}{2}(T_2 - T_1)(l_2 + l_1)$
 (3) $\frac{1}{2}(T_2 - T_1)(l_2 - l_1)$ (4) $\frac{1}{2}(T_2 + T_1)(l_2 + l_1)$
 (5) $\frac{1}{2}(T_2 + T_1)(l_2 - l_1)$

Elasticity

10

A wire being stretched by a force F_1 , has a length of l_1 . If the force is increased gradually from F_1 to F_2 , causing the the length to increase from l_1 to l_2 , what is the elastic energy stored in the wire during this process?

- (a) $(F_2 - F_1)(l_2 - l_1)$ (b) $\frac{1}{2}(F_2 - F_1)(l_2 - l_1)$ (c) $\frac{1}{2}(F_2 + F_1)(l_2 - l_1)$
 (d) $\frac{1}{4}(F_2 + F_1)(l_2 - l_1)$ (e) $\frac{1}{2}(F_2 + F_1)(l_2 + l_1)$

As asked in the question, when the applied force is increased from F_1 to F_2 (in the proportionality limit), the applied average force is $\frac{1}{2}(F_1 + F_2)$. The extension is $(l_2 - l_1)$. Therefore, the work done or the energy stored

in the wire (rod) is $\frac{1}{2} (F_1 + F_2) (l_2 - l_1)$ [mean force X extension]. The applied force should be slowly increased as it has to be done without the change of temperature (isothermally). If heat is generated, then part of the work done will be transformed into heat. When the force is increased from zero to a value of F , then the average force is $F/2$ [$1/2 (0+F)$]. The average force when the force is increased from $(F_1 + F_2)$ is $\frac{1}{2} (F_1 + F_2)$.

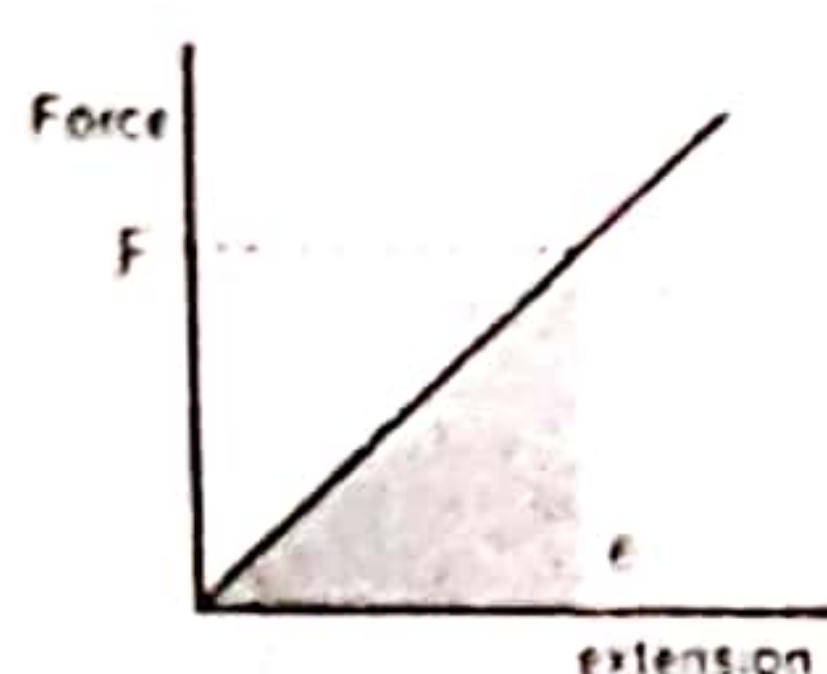


Figure 1(a)

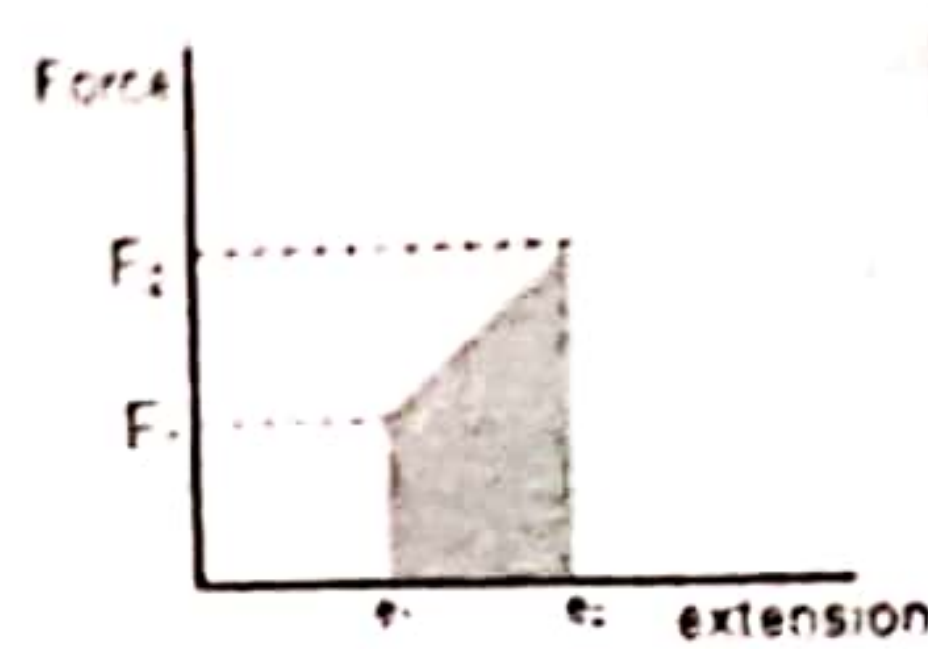
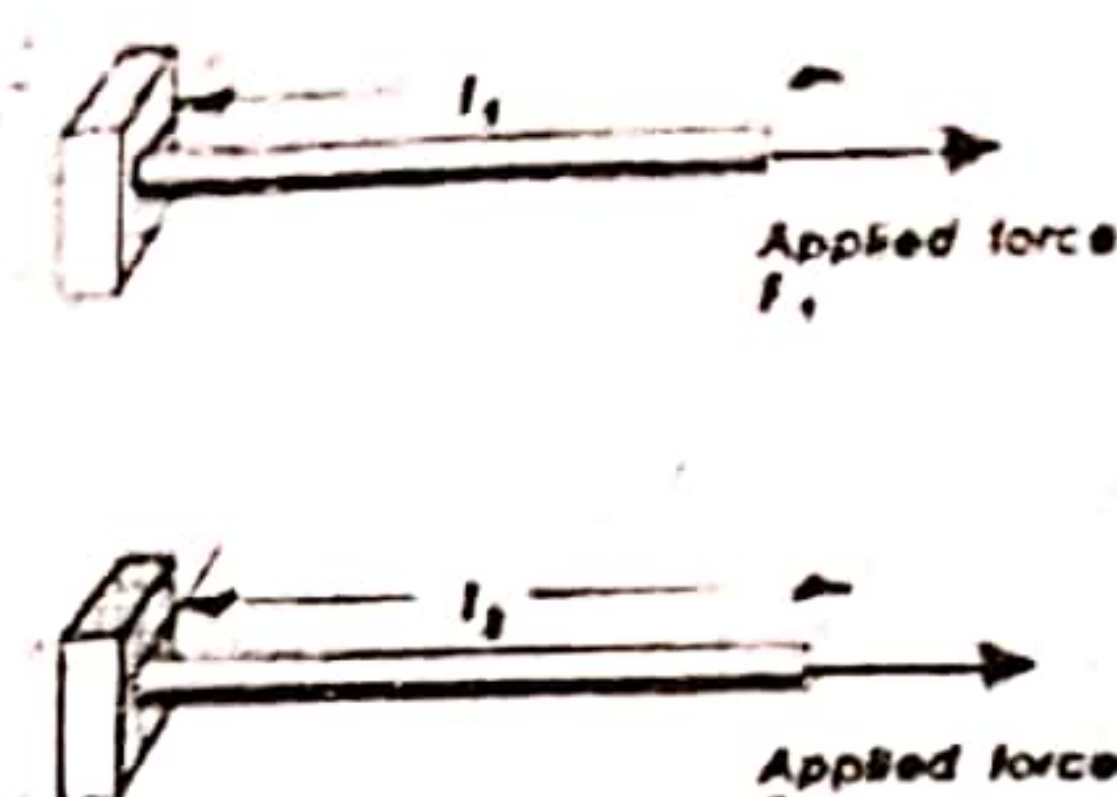


Figure 1(b)



15. Hydrogen gas in a container is maintained at standard temperature (300 K) and pressure ($1 \times 10^5 \text{ N m}^{-2}$). If the root mean square speed of hydrogen molecules is 2 km s^{-1} , what is the density of hydrogen in the container?

(1) 0.038 kg m^{-3} (2) 0.075 kg m^{-3} (3) 0.150 kg m^{-3} (4) 1.225 kg m^{-3} (5) 2.450 kg m^{-3}

04

Expansion of Gases

If you substitute to the equation $\sqrt{\overline{C^2}} = \sqrt{\frac{3P}{\rho}}$, you will get the answer easily. When talking about the pressure and the density, you need to use this equation. Actually, you do not need the temperature. The temperature is needed when P/ρ is substituted in T and M .

$$PV = m/M \cdot RT; P = m/V \cdot RT/M; P = \rho RT/M; P/\rho = RT/M$$

There are many questions that have been asked from this equation. Even the temperature is not needed, 300 K means the room temperature that we normally consider. The standard temperature is 273 K (0°C). $\overline{C^2} = \frac{3P}{\rho}$; $\rho = \frac{3 \times 10^5}{(2 \times 10^3)^2} = \frac{3}{4} \times 10^{-1} = 0.075 \text{ kg m}^{-3}$

Normally, such questions are given to find $\overline{C^2}$ when P and ρ are given.

16. A composite rod is formed by connecting two rods A and B as shown in the figure. Longitudinal wave velocities in rods A and B are 3210 m s^{-1} and 6420 m s^{-1} , respectively. A longitudinal pulse applied at the free end of the rod A propagates with a wavelength of 2 m. What is the wavelength of this wave when it propagates through rod B?

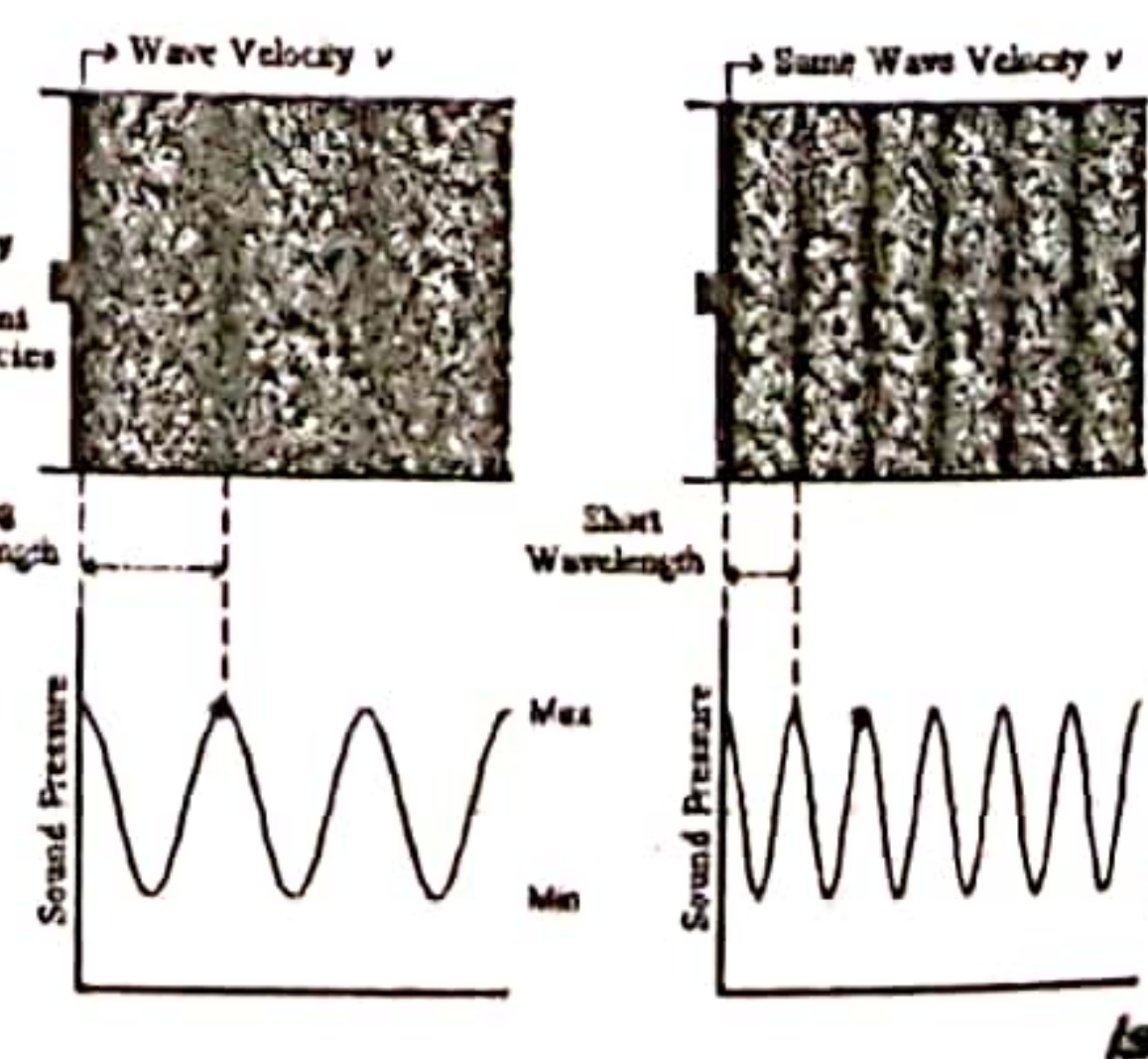


(1) 1 m (2) 2 m (3) 3 m (4) 4 m (5) 5 m

03

Longitudinal Waves

When a wave is going from one medium to another medium, its frequency does not change. According to the speed of the wave, the wavelengths are adjusted. Therefore, it should be $3210/2 = 6420/\lambda$ as $v_1/\lambda_1 = v_2/\lambda_2$. As the speed is getting double, then the wavelength should also be doubled. That means $\lambda = 4\text{m}$. These questions are very easy. I feel that it is better if I could put some questions with many sentences that are there before the question 13, to later questions. Even if we think that they are easy, the students spend more time on questions with many sentences.



17. The magnitude and the direction of the electric field at point A due to the point charge distribution shown in the figure are

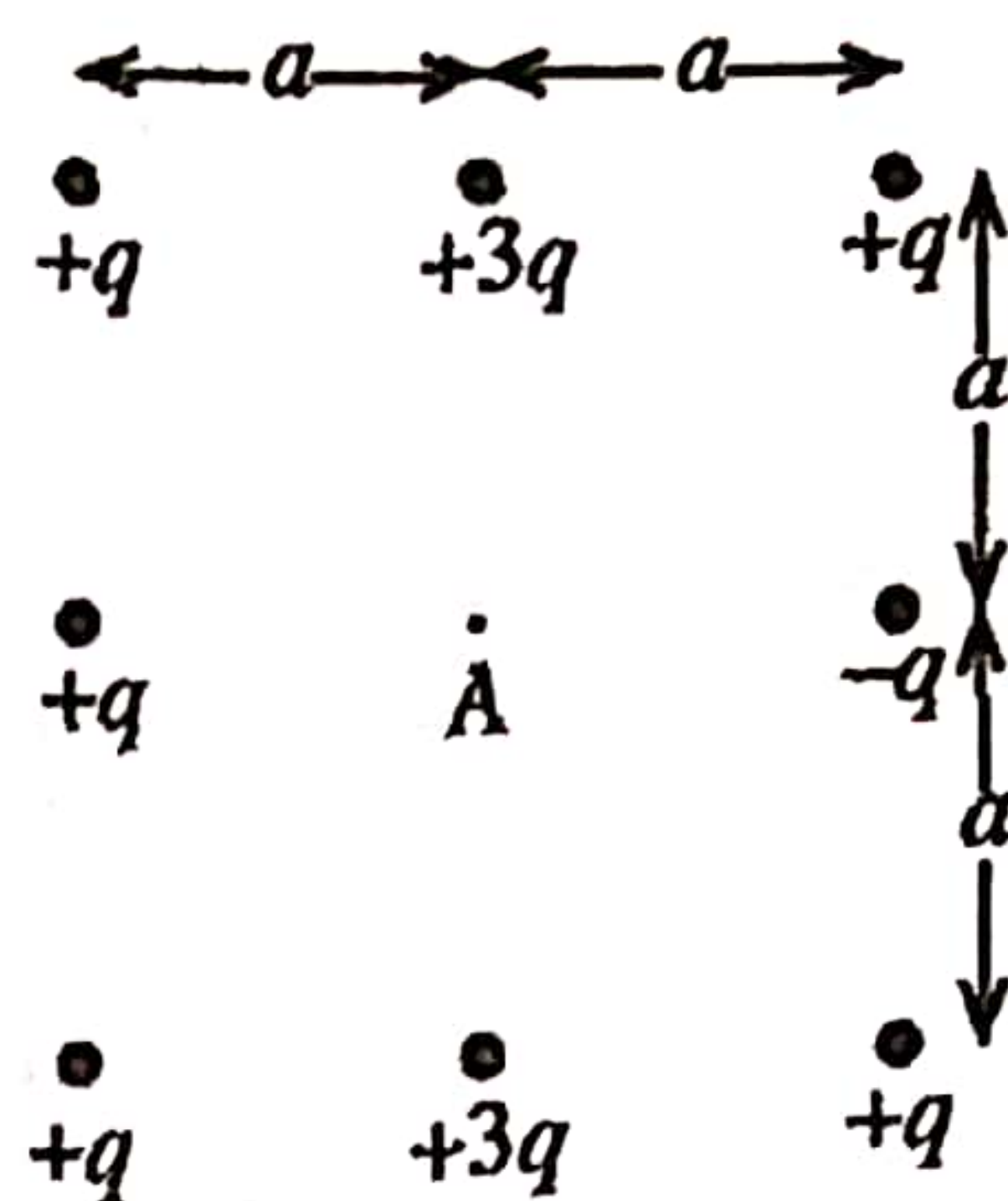
(1) $\frac{2q}{4\pi\epsilon_0 a^2} \rightarrow$

(3) $\frac{2q}{4\pi\epsilon_0 a^2} \leftarrow$

(5) $\frac{6q}{4\pi\epsilon_0 a^2} \downarrow$

(2) $\frac{q}{4\pi\epsilon_0 a^2} \uparrow$

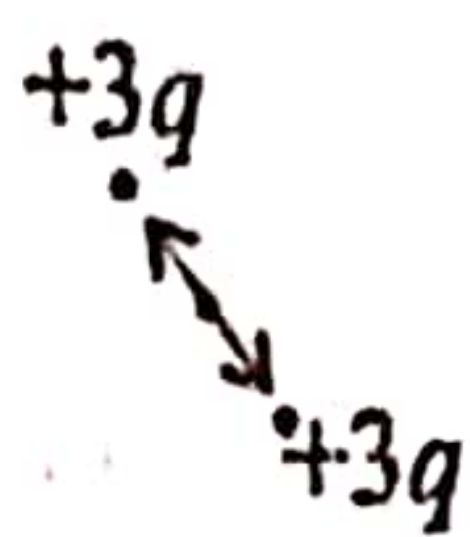
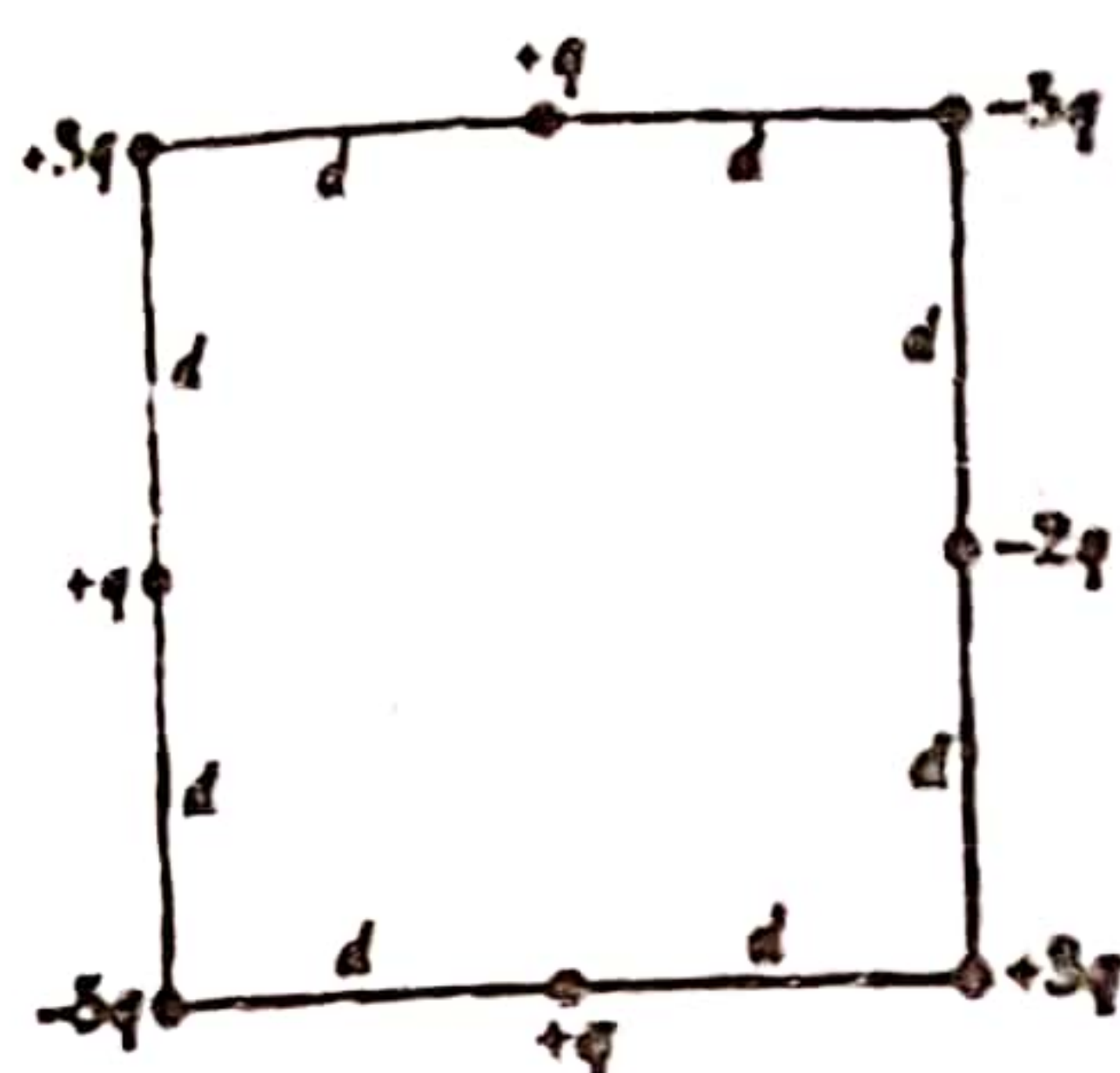
(4) $\frac{6q}{4\pi\epsilon_0 a^2} \uparrow$



Electrostatic Force Field

06

Consider the given eight point-charges in the question.



How do you find the magnitude and the direction of the electric field at the centre? The centre is in the middle of the square. The electric field intensity from the charge couples (+ or -) that are equal in

magnitude but are kept in the opposite direction equidistant to the centre is zero. To find the electric field intensity, keep a unit positive charge at the centre. The +3q charge at the left corner repels the unit charge.

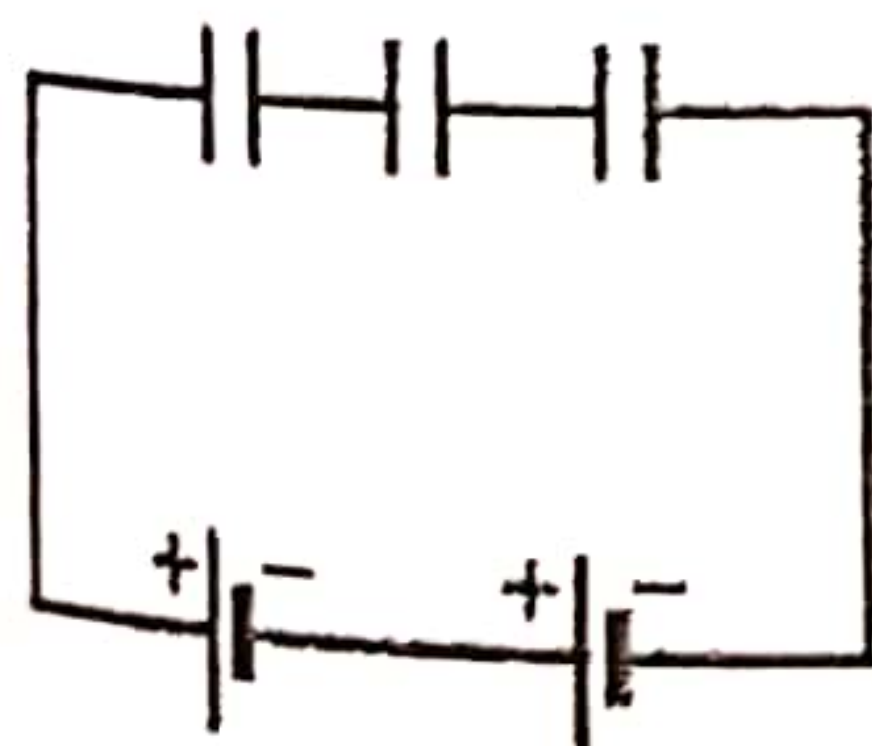
Even the +3q charge at the right corner also creates a repulsive force which is equal but opposite in direction. So, they are being cancelled off with each other. Therefore, you can neglect all the charges of same sign that are equidistant but in opposite direction. Do not care about them. Finally, you will be left with +q and -2q. Think that we kept a -q charge instead of -2q.

Now by +q, the unit positive charge is being repelled. By -q, it is being attracted. Both are acting to the same side. Therefore, E at the centre = $2 \cdot \frac{1}{4\pi\epsilon_0} \frac{q}{a^2}$. The direction is towards \rightarrow . If this is the only answer to this direction, then the selection is very easy.

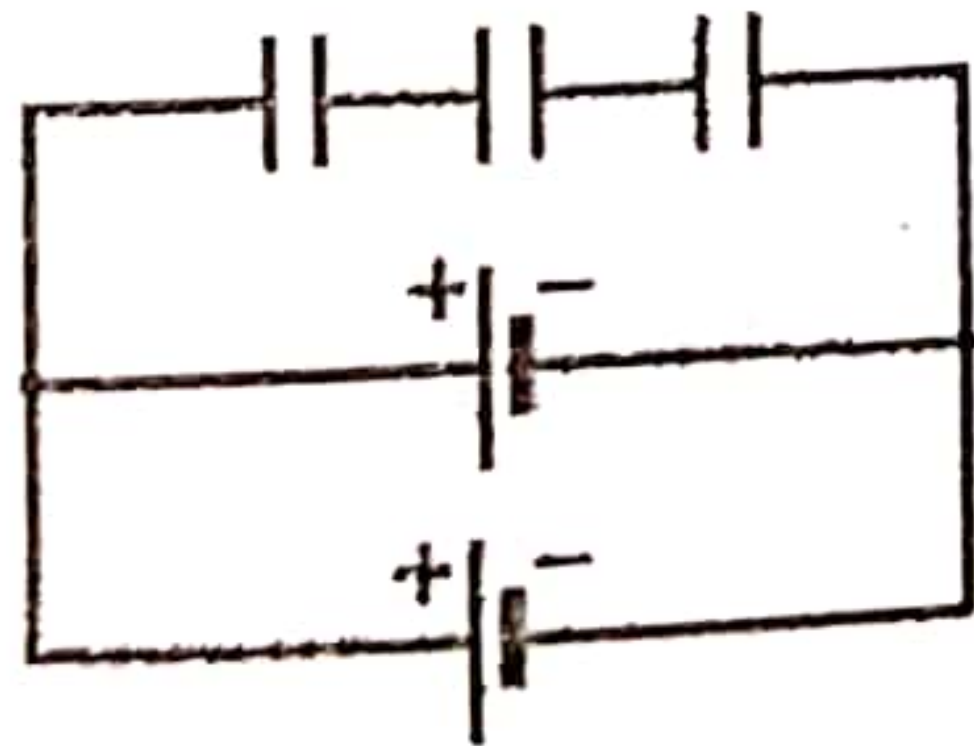
Males and males, as well as females and females are repelled by each other. Therefore, a net field is not created in the middle. Normally, this is what happens. There can be people who are abnormal.

Males repel with each other and get attracted by a female. Therefore, neglect all males and females that are equidistant. Just only consider the equidistant male and female.

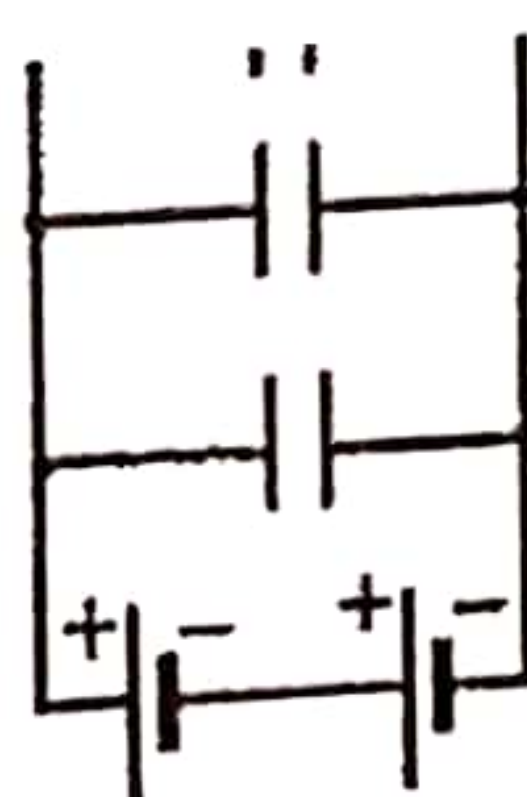
18. Three capacitors with equal capacitance and two batteries with equal electromotive force (emf) are given to construct a circuit to store energy. Which of the following circuits stores the maximum energy?



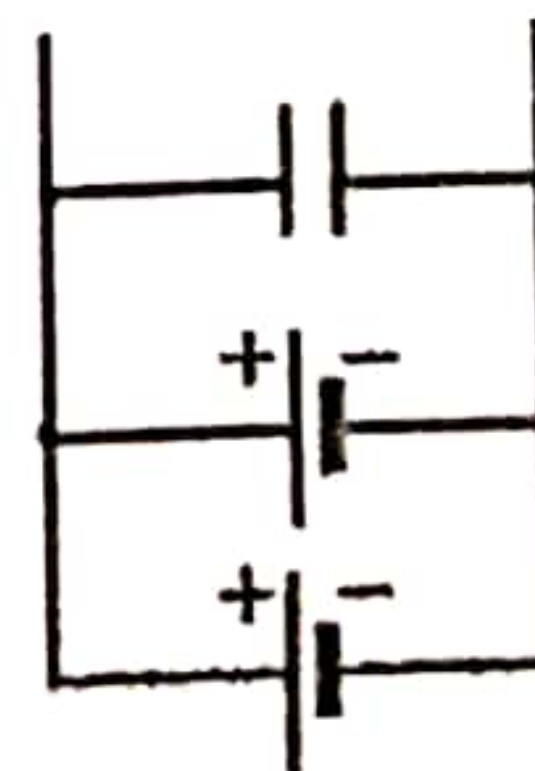
(1)



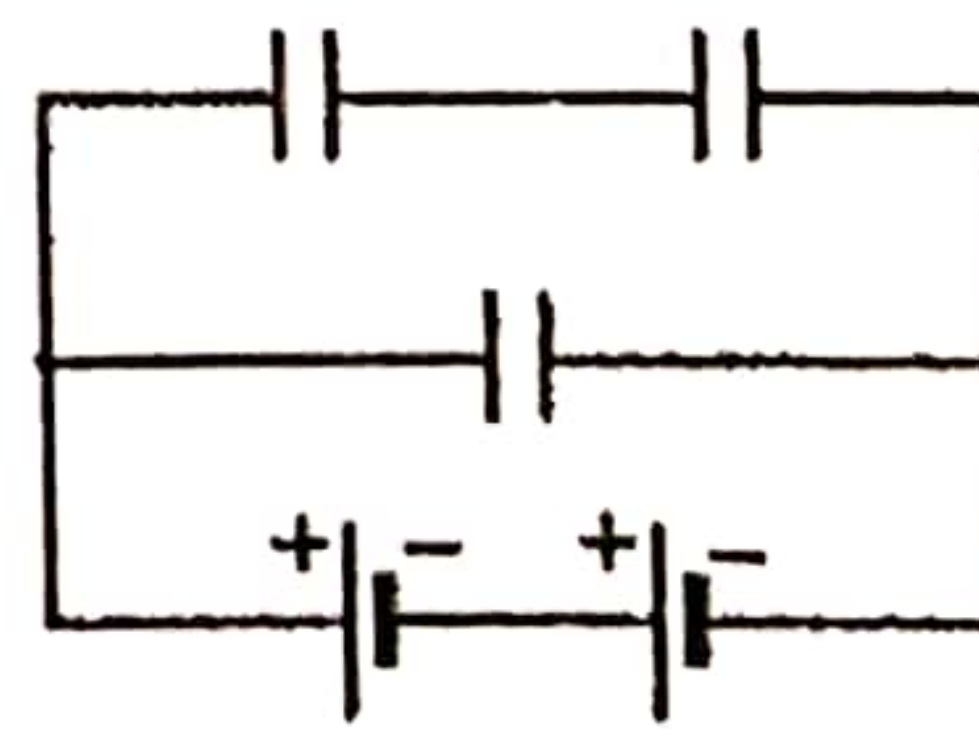
(2)



(3)



(4)



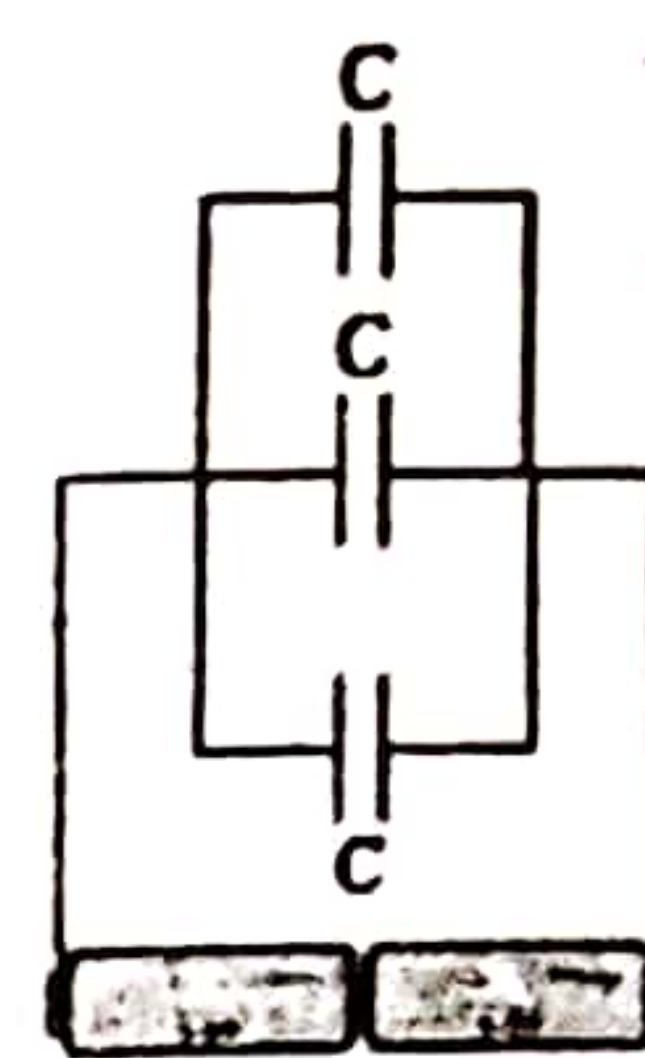
(5)

Capacitance and Capacitors

06

Equivalent capacitance is maximum when the capacitors are connected in parallel. The maximum e. m. f of the cells can be obtained when they are connected in series. Therefore, to get maximum capacitance as well as maximum e. m. f, definitely you need to connect the three capacitors in parallel and the two cells in series. If the capacitance of a capacitor is C , then the equivalent resistance is $3C$. From any other arrangement you cannot get $3C$. If the e. m. f of a cell is E , then when the cells are connected in series the net e. m. f is $2E$. The stored energy of the shown correct arrangement is $\frac{1}{2} \times 3C \times (2E)^2$.

$$C_{eq} = C_1 + C_2 + C_3 \times$$



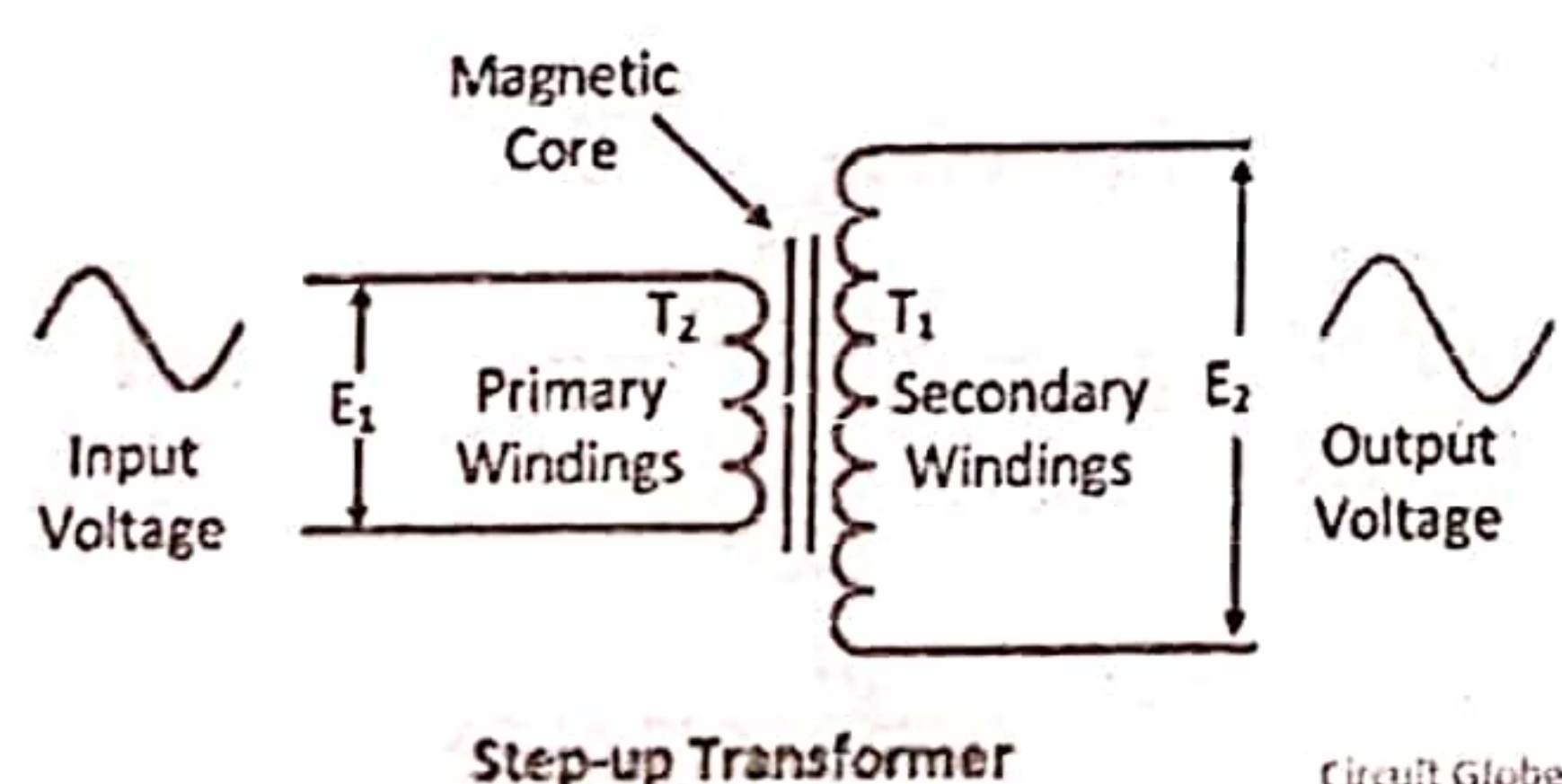
19. When a current of 6A is flowing through the primary coil of an ideal transformer having power of 60 W, output voltage is 12 V. Select the correct answer which gives the type of the transformer and the current ratio (Primary current : Secondary current).

- (1) Step down and 6 : 5 (2) Step down and 5 : 6 (3) Step up and 1 : 2
(4) Step up and 5 : 6 (5) Step up and 6 : 5

08

Mutual Induction

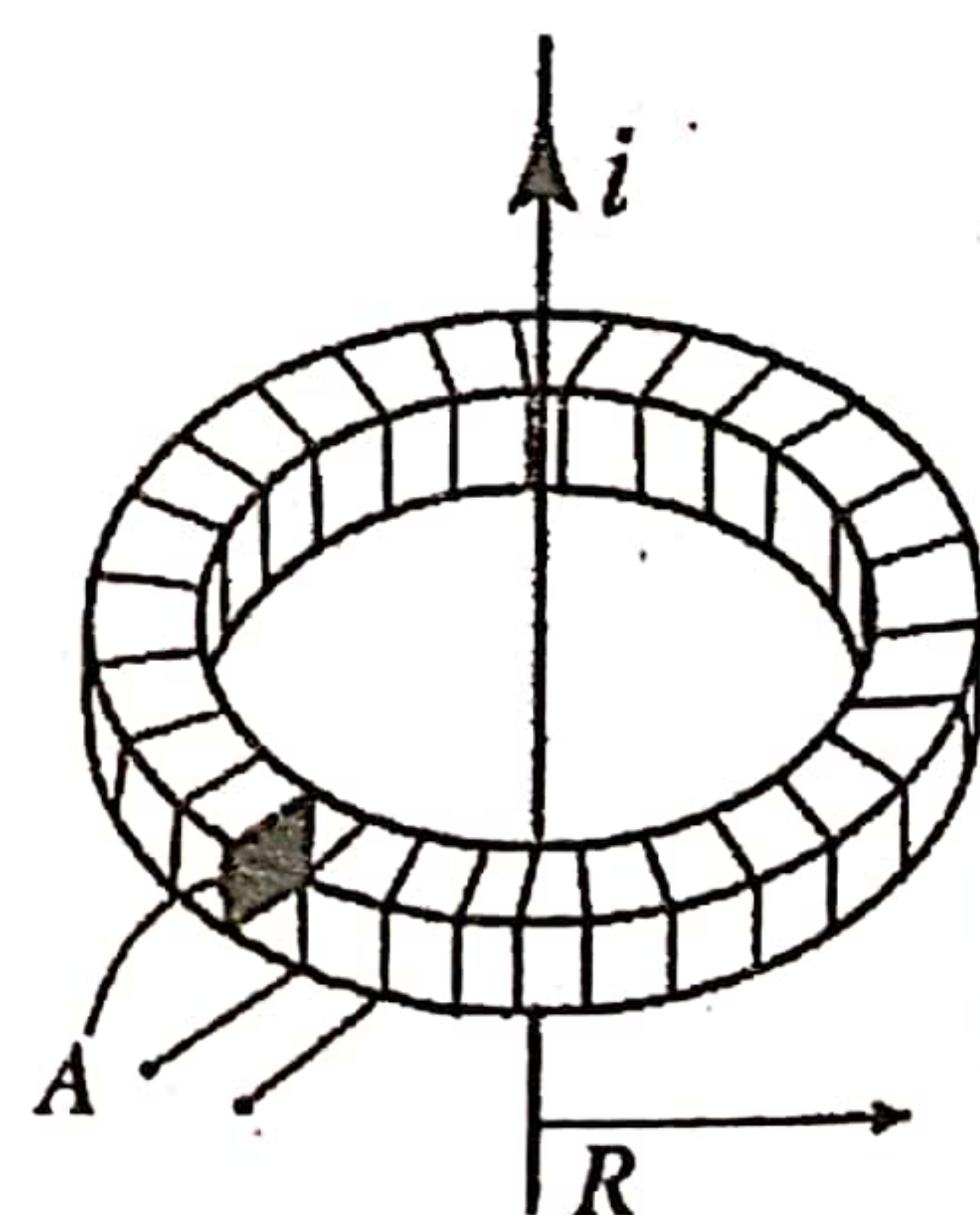
A step-up transformer has been shown in the figure.



If the power is 60 W and the current through the primary coil is 6 A, then the primary voltage is $60/6 = 10$ V ($W = Vi$). The secondary voltage (output) is 12 V. This is a step-up transformer ($12 > 10$). Normally, we consider such transformers as ideal. That means power of 60 W does not change. Therefore, secondary current $= 60/12 = 5$ A. In the step-up transformer, primary current/ secondary current $= 6/5$.

20. A coil is made by winding N number of turns around a plastic ring of mean radius R and cross sectional area A as shown in the figure. This coil is placed coaxially with a long straight wire carrying a current i . If the rate of change of current through the straight wire is $i_0 \cos \omega t$, which of the following expressions gives the electromotive force (emf) induced in the coil?

- (1) $\mu_0 AN i_0 \cos \omega t$ (2) $\mu_0 AN^2 i_0 \sin \omega t$
(3) $\frac{\mu_0 AN}{\omega} i_0 \sin \omega t$ (4) $\frac{\mu_0 AN}{2\pi R} i_0 \cos \omega t$
(5) $\frac{\mu_0 AN}{4\pi^2 R^2} i_0 \cos \omega t$



07

Magnetic Effect of Electric Current

A rigid circular loop of wire of radius R and N turns is placed inside a long solenoid having N_0 turns per metre and carrying current $I(t) = I_0 \sin \omega t$. The centre of the loop lies on the axis of the solenoid and the plane of the loop is normal to the axis. The induced electrical field E_i at any point on the loop is given by,

a) $E_t = 0$ b) $E_t = -\frac{\mu_0 I_0}{2} \cos \omega t$ c) $E_t = -\frac{\mu_0 I_0 N N_0}{2} \cos \omega t$ d) $E_t = -\frac{\mu_0 I_0 N N_0 r \omega}{2} \cos \omega t$

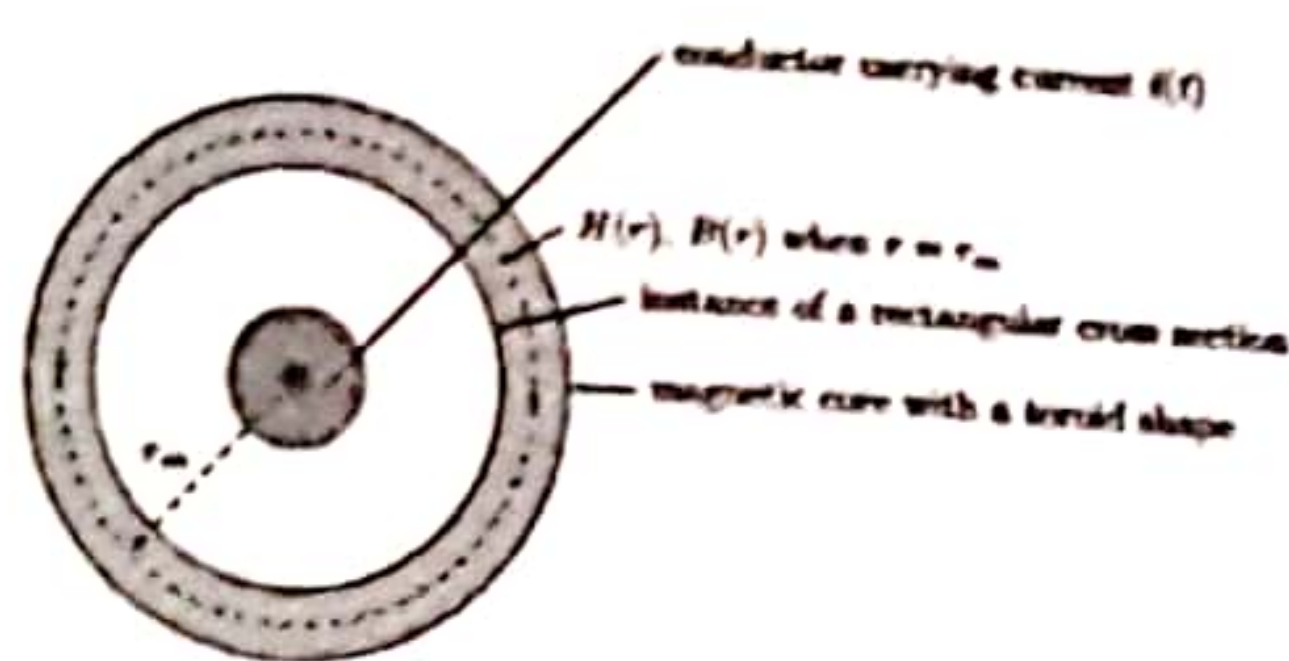
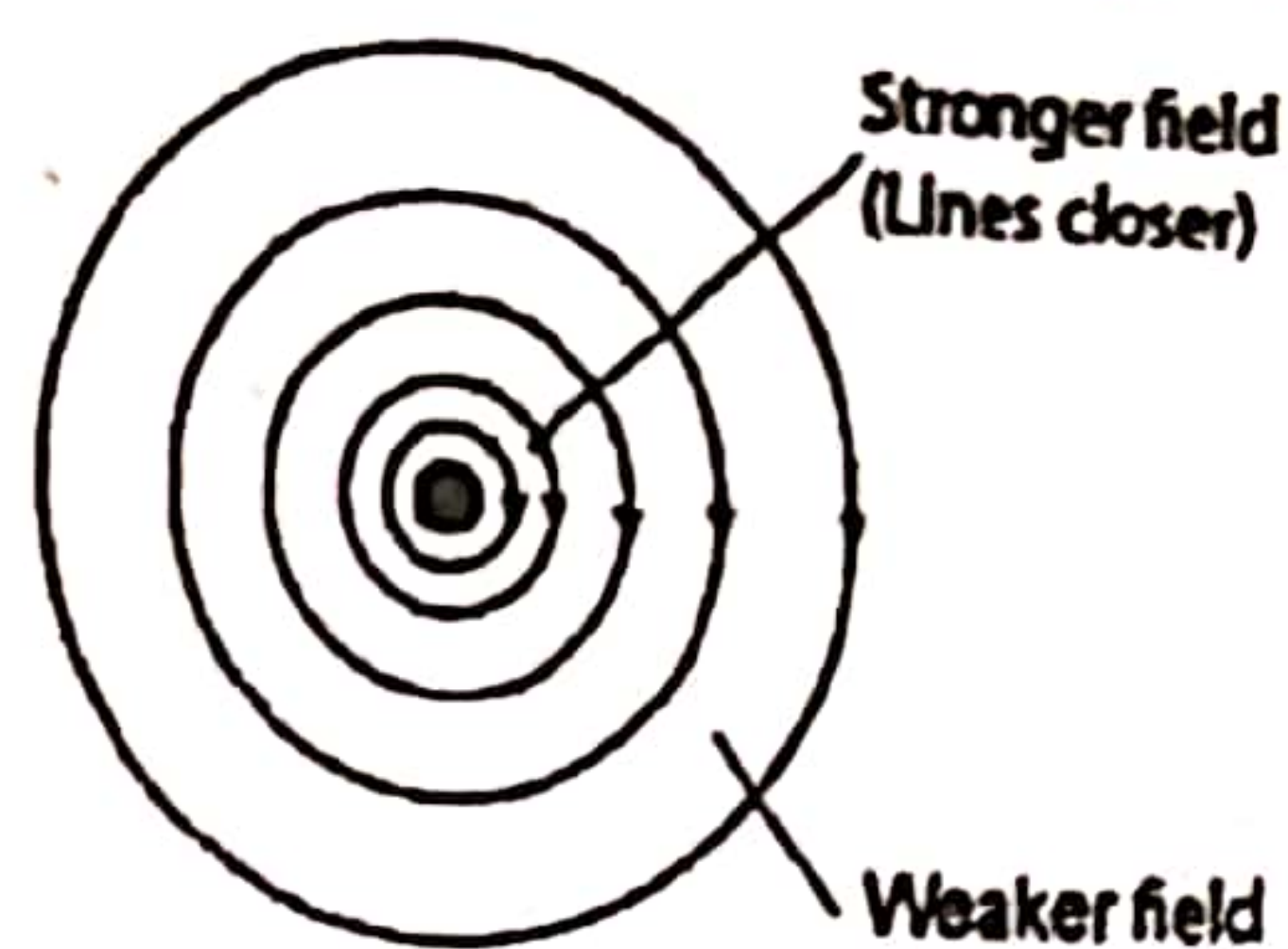
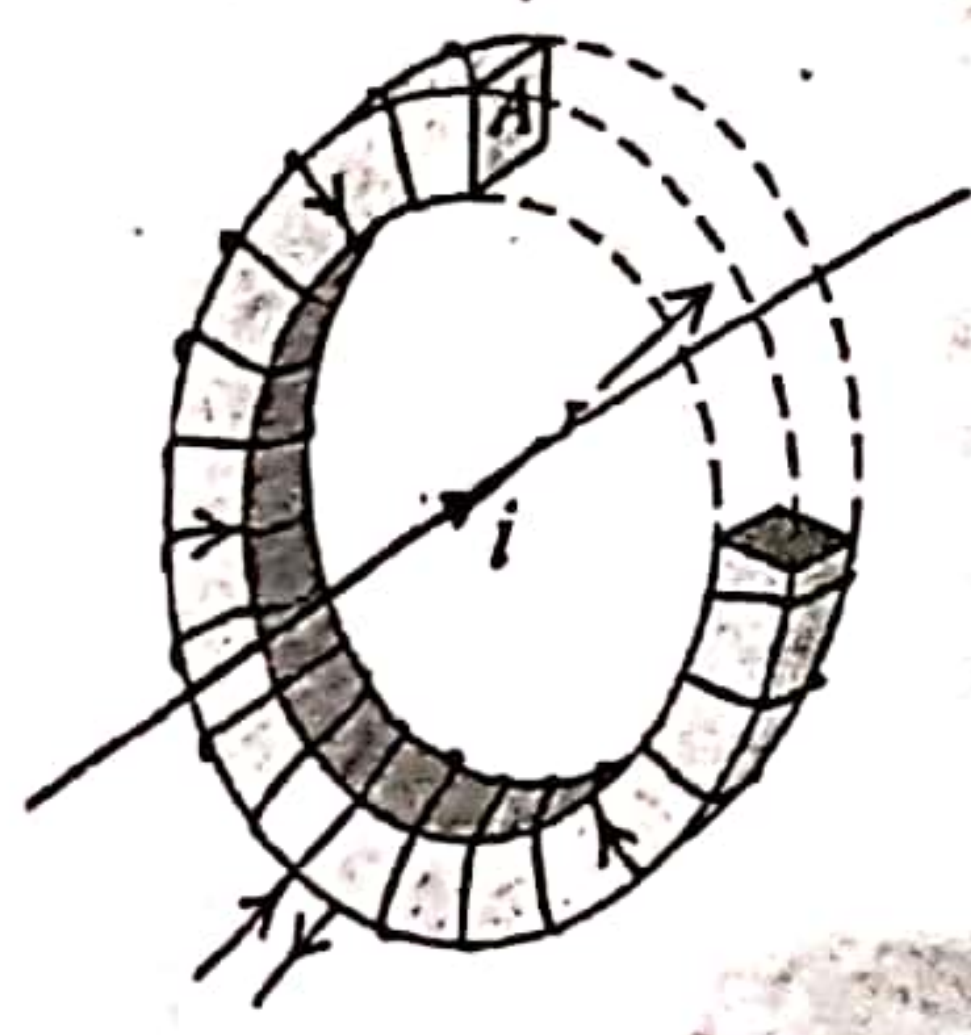


Figure 1: Frontal view of a toroid surrounding a current-carrying conductor. By convention, instantaneous positive current is assumed flowing out of the paper, hence the magnetic field is depicted in the counter-clockwise direction per the right-hand rule.

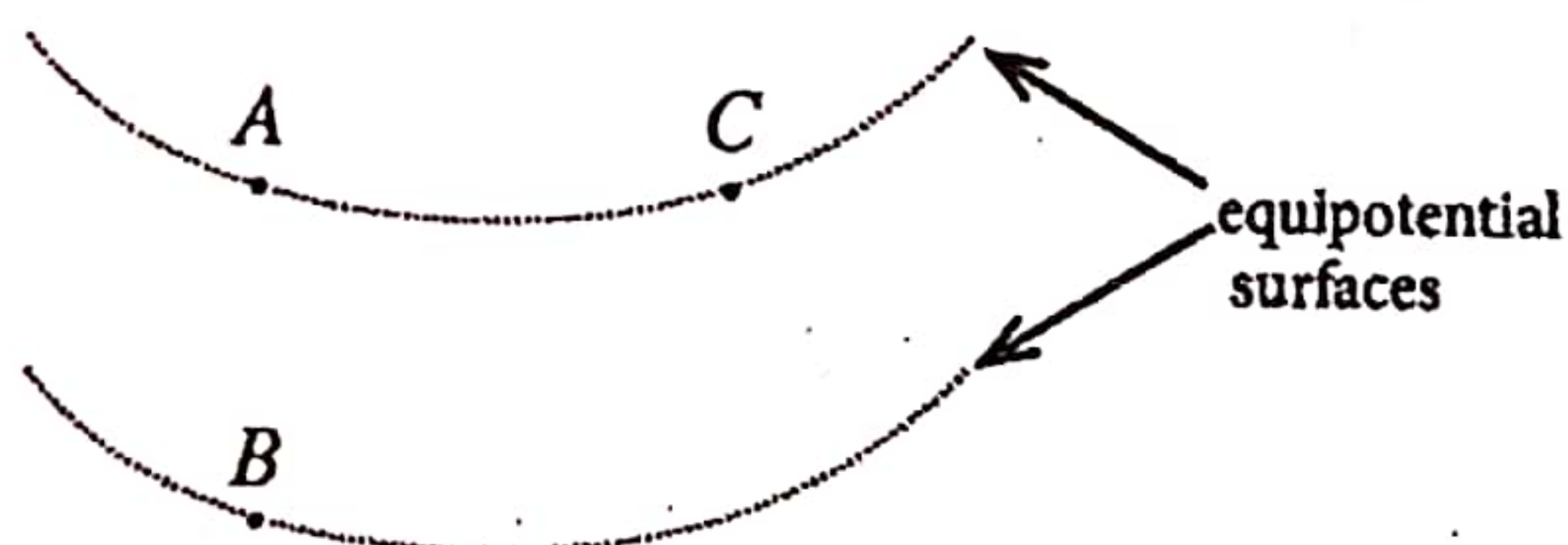


We know that the magnetic flux density B of a straight long wire with current in a distance R is $\mu_0 i/2\pi R$. Now this wire is kept perpendicularly according to the figure along the axis of a coil with a soft iron core of mean radius R and cross-sectional area A and N turns. If we consider the magnetic flux density inside the coil is uniform, then the magnetic flux across the coil $= BAN = \mu_0 ANi/2\pi R$ [magnetic flux density \times perpendicular area \times number of turns]. Now if the current in the wire is varying in a rate of i' then the rate of magnetic flux change is $\mu_0 ANi'/2\pi R$. According to Faraday's law, this is the magnitude of the induced e. m. f of the coil.

If $i' = i_0 \cos \omega t$, then the induced e. m. f is $(\mu_0 AN/2\pi R) i_0 \cos \omega t$. If the current is considered as an alternating current, then the normal tradition is to write as. The rate in which this current varies with time $di/dt = i_0 \omega \cos \omega t$. [The students who are in the biology stream do not look at this matter.] Therefore, the current which is given in the paper is not the peak current. To tally with dimensions, it should be $i_0 \omega$. But one can argue that what is given in i_0 is $i_0 \omega$.



21. Consider the points A , B , and C on two equipotential surfaces as shown in the figure. When a proton moves from A to B , the electric field does a work of $3.2 \times 10^{-19} \text{ J}$ on it. Charge of an electron is $-1.6 \times 10^{-19} \text{ C}$. The electric potential differences V_{AB} , V_{BC} , and V_{CA} , respectively, are
- (1) 2 V , -2 V , and 0 V
 - (2) 2 V , -2 V , and 2 V
 - (3) -2 V , 2 V , and 0 V
 - (4) 0.5 V , -0.5 V , and 0 V
 - (5) -0.5 V , 0.5 V , and 0 V



Electrostatic Potential

06

The electric field lines and the equipotential surfaces from a point positive charge is shown here. When taking a positive charge from the equipotential surface A to a point in B , the electric field is doing work on it. Positive and positive repel each other. It is clear that the potential of the equipotential surface in point A is greater than the potential of equipotential surface in point B . The potential difference between two equipotential surfaces is equal to the work done by the electric field when moving a unit positive charge. If a proton is moving, then the charge of a proton is $+1.6 \times 10^{-19} \text{ C}$. If the work done by the field when a proton is moving is $3.2 \times 10^{-19} \text{ J}$, then when a unit charge ($+1 \text{ C}$) is moved, the work done it will be 2 V .



[$3.2 \times 10^{-19} / 1.6 \times 10^{-19} = 2 \text{ V}$]. So, $V_{AB} = +2 \text{ V}$. V_{AB} means $V_A - V_B$. As points A and C are in the same equipotential surface, $V_{AC} = V_{CA} = 0$. If $V_{AB} = +2 \text{ V}$, then V_{BA} and V_{BC} should be -2 V . $V_{AB} = +2 \text{ V}$; $V_{BC} = -2 \text{ V}$; $V_{CA} = 0$

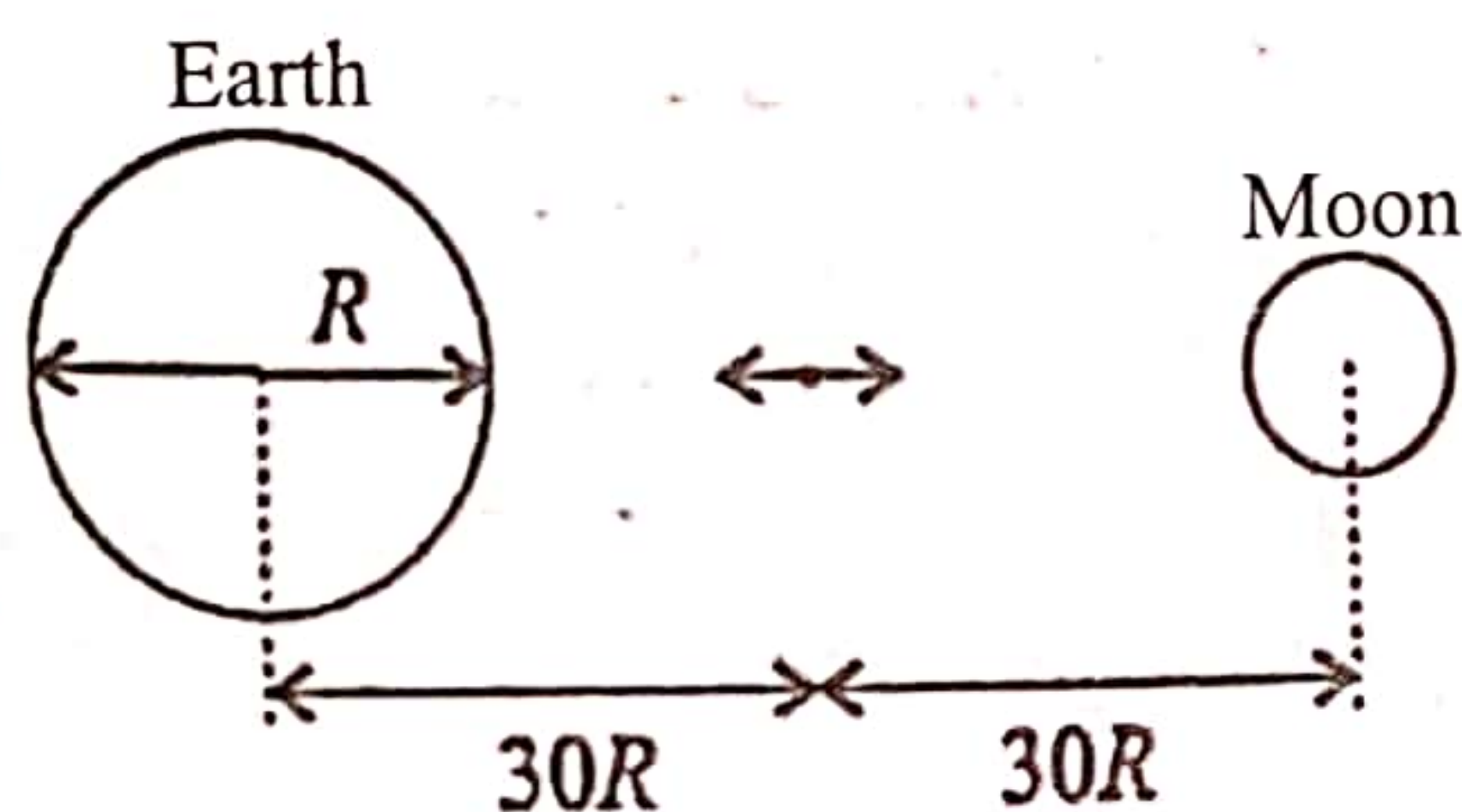
22. A celestial object is located at the mid point of the line joining the centres of the earth and the moon at a certain time. The mass of the moon is 0.0123 times the mass of the earth. Assume that the distance between the centres of the moon and the earth is 60 times the radius of the earth. The acceleration of the object due to the gravity of both the earth and the moon in terms of g is approximately,

(1) $1.1 \times 10^{-6} g$ (2) $1.1 \times 10^{-3} g$ (3) $3.3 \times 10^{-2} g$ (4) $0.5 g$ (5) $1.0 g$

05

Gravitational Force Filed

Look at the shown instance.



Consider an object which is placed in between the line joining the centres of the earth and the moon. If the mass of the moon is 0.0123 times of the mass of earth, then find the acceleration of the object in g (the gravitational acceleration on earth's surface).

The acceleration on the object due to earth = $GM_E / 30^2 R^2$

The acceleration on the object due to moon = $GM_M / 30^2 R^2$

Net acceleration = $GM_E / R^2 [1/30^2 - 0.0123/30^2]$

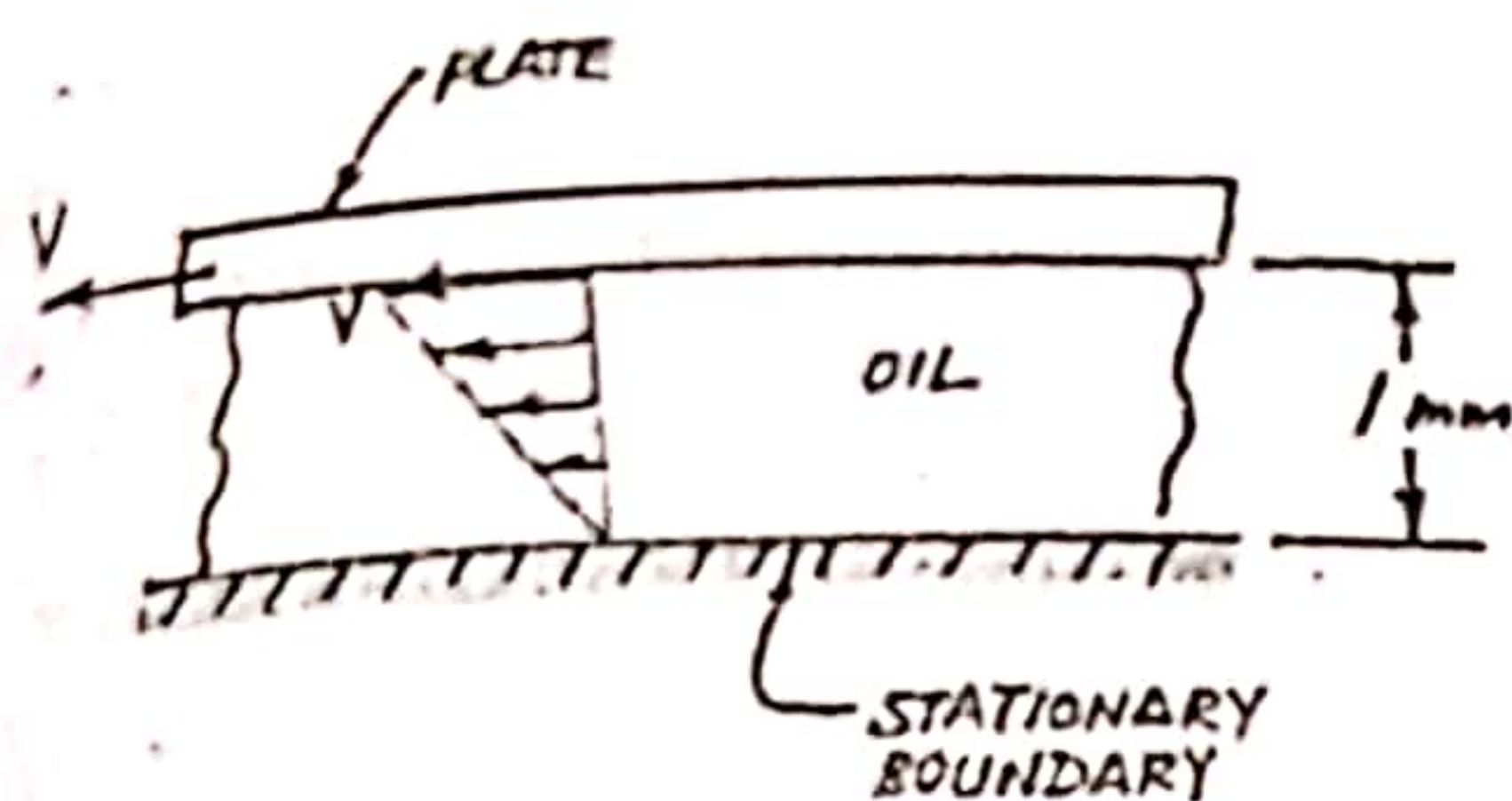
Write this expression directly in your rough worksheet. GM_E / R^2 is the acceleration of an object which is kept on the earth's surface which we consider as g .

So, the answer = $g (0.9877 / 900) = 1.1 \times 10^{-3} g$ (if it is estimated to the first decimal point)

There is a simplification. It takes some time. As 10^{-3} is in one answer, you can say that the correct answer value cannot be cared. But when answering, the children does not think of this issue. They try to simplify till the last end.

Another fact is that if we do not consider the moon, then the acceleration only from the earth is near to the first decimal point is taking this value. $g/30^2 = 1.1 \times 10^{-3} g$. Therefore, if we did not consider the moon and ask the acceleration at a point that is in $30R$ distance from the earth, the children could have saved some time. Even the before and after zeros of the decimal in the value 0.0123 are not successful numbers. Here the second decimal (which is 1) is not 5 or higher than 5. Therefore, this value does not affect for the first decimal place of the answer. I feel that the question could have been simpler if it was asked considering only about the earth. When $1/1000$, you will get 1×10^{-3} . When $1/900$, then the answer must be little more than 1×10^{-3} . The only answer which has greater than 1 with 10^{-3} is 1.1×10^{-3} .

23. The gap of 2 cm between two horizontal plates of surface area 500 cm^2 is filled with an oil having the coefficient of viscosity 0.2 N s m^{-2} . A horizontal force of 5 N is applied to the upper plate while keeping the lower plate at rest. What is the velocity of the middle layer of the oil if the velocities of the oil layers vary linearly across the gap between the plates?
- (1) 2.5 ms^{-1} (2) 5 ms^{-1} (3) 10 ms^{-1} (4) 25 ms^{-1} (5) 50 ms^{-1}



Viscosity

10

The questions with oil layers among the plates can be seen in many past papers.

$5 = 0.2 \times 500 \times 10^{-4} \times [(v-0)/2 \times 10^{-2}]$ where v is the velocity of the oil layer which touches with the upper plate. There are numbers which can be easily simplified. 5 and 5 are cut off. $0.2/2 = 0.1$.

$$v = 10^{-2}/10^{-3} = 10 \text{ ms}^{-1}.$$

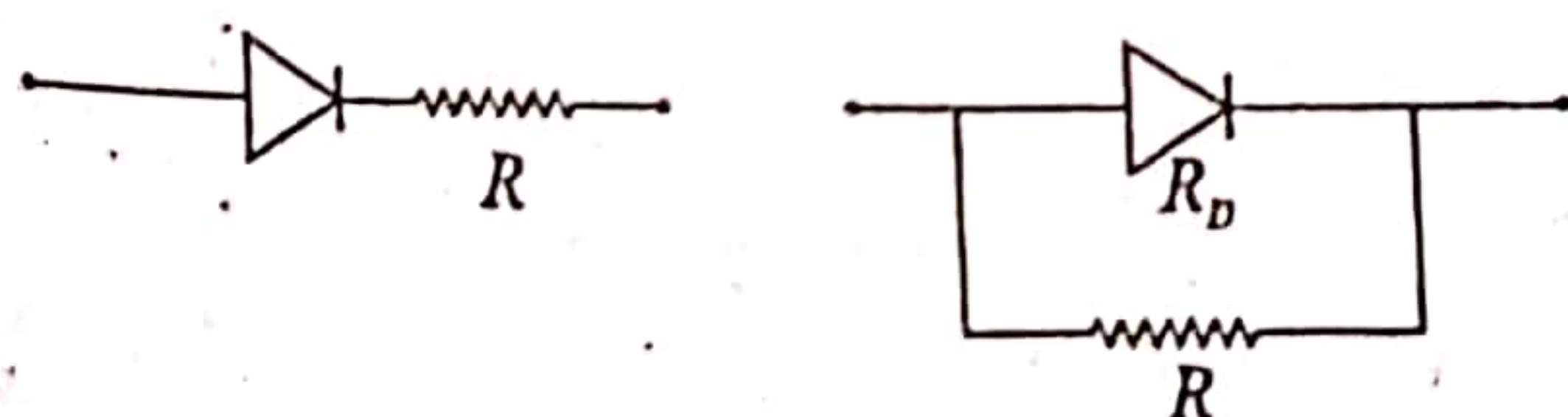
As the velocity change is uniform, the velocity of the middle layer is 5 ms^{-1} .

24. A diode and a resistor are connected in such a way that only two terminals are available for external connections. When a voltage of 1 V is applied across the external terminals, the current through the circuit is 50 mA. When the applied voltage is reversed, the current doubles. What are the forward bias resistance of the diode, and the value of the resistor?

	Resistance (Ω)	
	Diode	Resistor
(1)	0	20
(2)	10	10
(3)	10	20
(4)	20	10
(5)	20	20

Semi Conductor Diodes

09



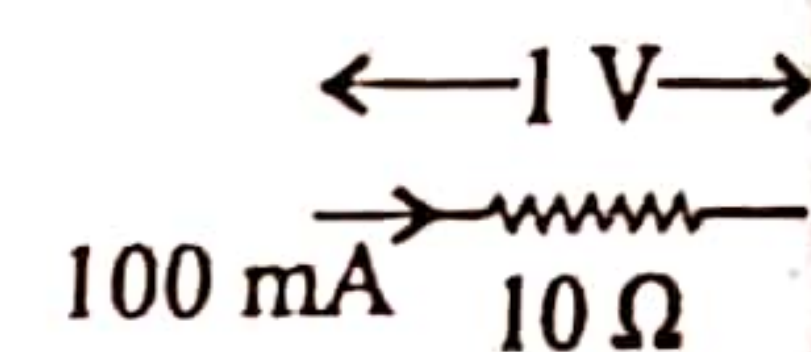
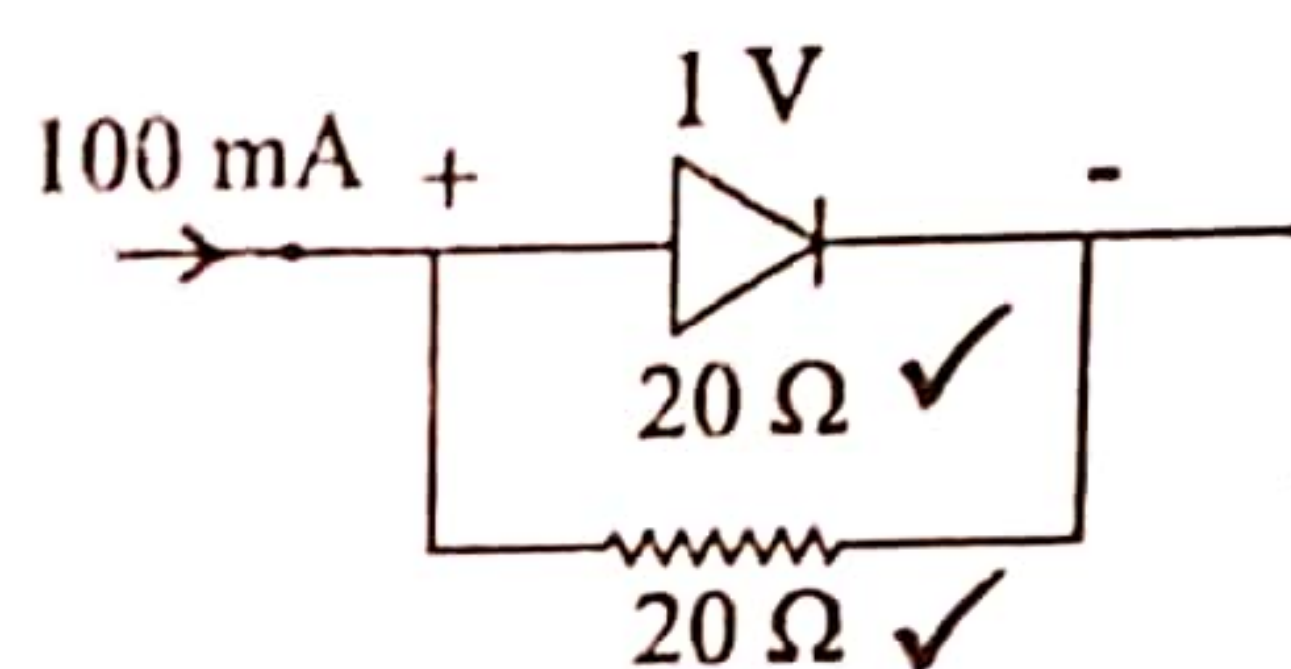
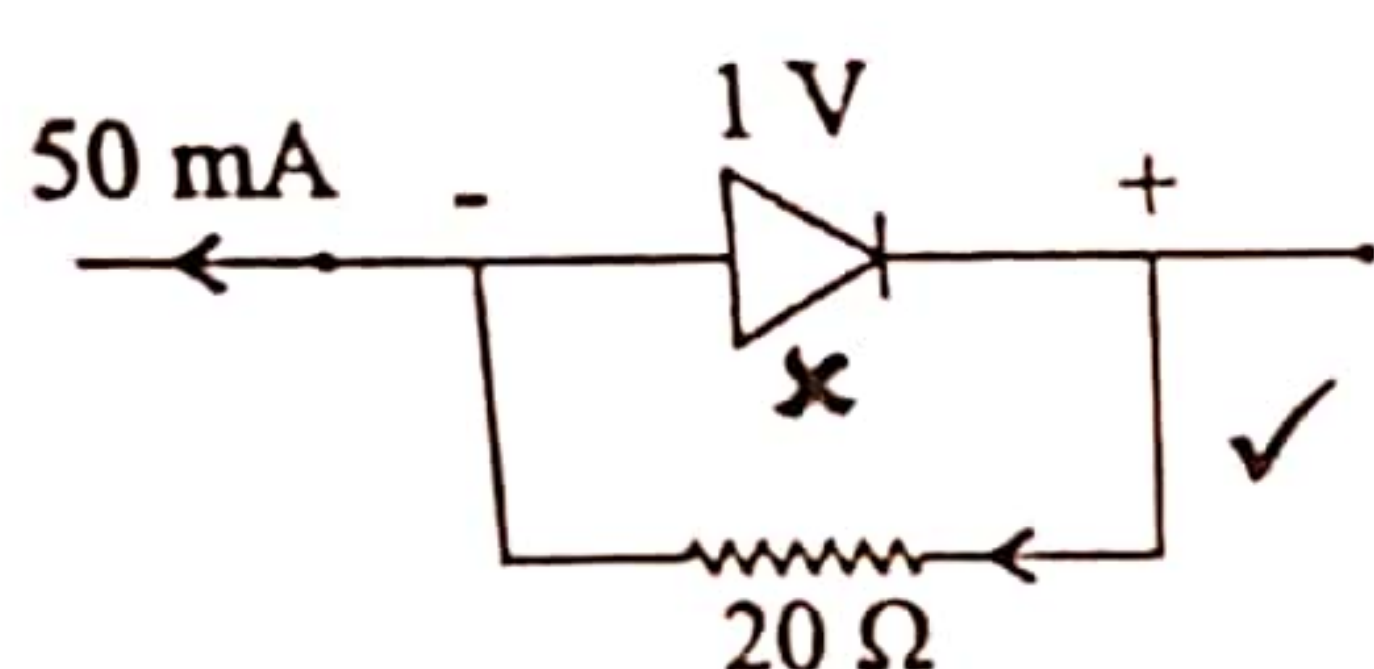
According to the figures, there are two ways that a diode can be connected with a resistor.

That is by in series and in parallel. First you need to decide that whether the resistor is connected in series or parallel. Once you read the question, you can decide that it has been connected in parallel. Why? If it was connected in series, then the current that is flowing across the resistor, as well as the diode that means across the circuit should be zero as the diode gets reverse biased each occasion. There is no such a mentioned fact in the question. Therefore, it is clear that the diode and the resistor is connected in parallel. Next, from the given two instances, you need to think whether each instance is relevant with the forward or reverse biased mode of the diode.

Initially, when 1 V is applied, the current across the circuit is 50 mA whereas when the direction of the voltage was changed, it has been mentioned that the current gets double value of 100 mA. When the diode is forward biased, the current flows across the diode as well as the resistor. When the diode is reverse biased, the current flows across the resistor only. We know that when two resistors are in parallel, the equivalent

resistance of them is lesser than the lowest resistor of the two. Therefore, more current is flow when the resistors are in parallel. That means the first instance is relevant to the reverse biased mode of the diode. Then there is no current across the diode. 50 mA is flow across R only. Therefore, according to $V = iR$, $1 = 50 \times 10^{-3} R$. $R = 20 \Omega$.

Next, when the diode is forward biased, there is a current flow across the diode as well as from the resistor. If the current is getting doubled, then the resistance should be halved (from 20). So, the forward biased resistance of the diode should be 20Ω . When 20Ω and 20Ω are parallel, the equivalent resistance gets 10Ω . When the equivalent resistance is halved, then the total current flow will be doubled (for a constant voltage). If you understand it properly, then you can solve this question without any rough worksheet. Initially you need to realize that the instance where the current is not flowing has been given. The current is increased when it flows across the parallel arrangement.



25. An air mass in a cyclone moves around its eye in a spiral path as shown in the figure. The velocity of the air mass at a radial distance of 80 km from the centre of the eye is 150 km h^{-1} . What could be the velocity of the same air mass at a radial distance of 40 km from the centre of the eye?

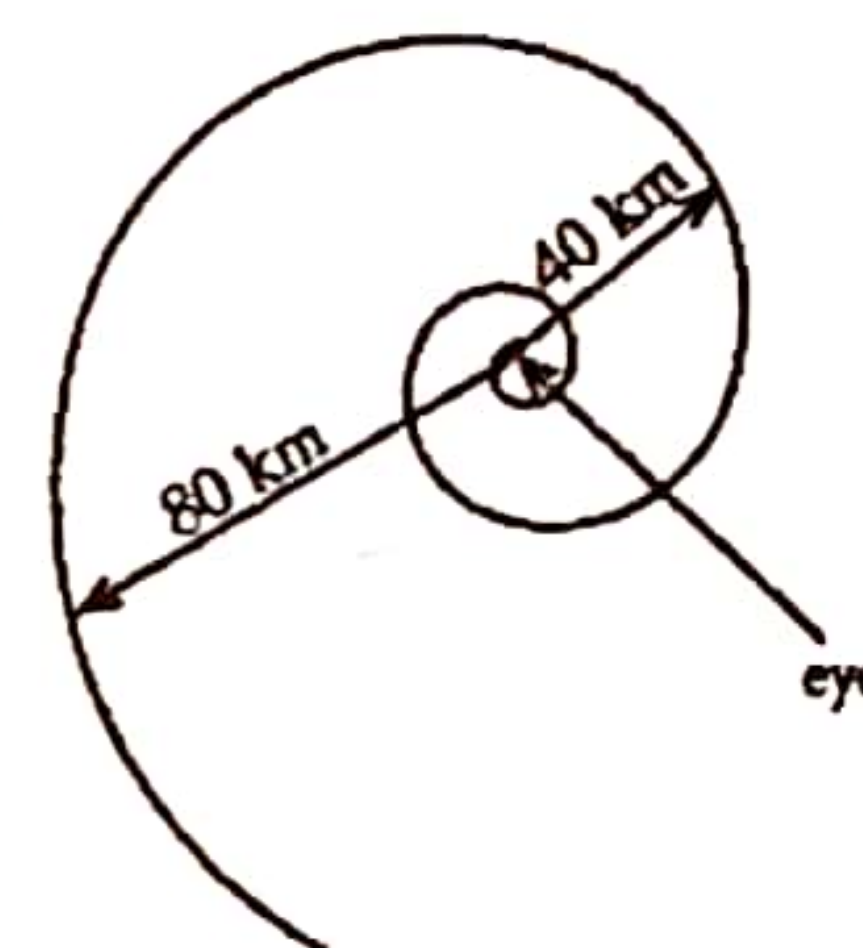
(1) 75 km h^{-1}

(2) 150 km h^{-1}

(3) $150\sqrt{2} \text{ km h}^{-1}$

(4) 300 km h^{-1}

(5) 450 km h^{-1}



02

Work Power and Energy

You need to decide that application of conservation of angular momentum as soon as you see the question. There is no other principle which can be applied. When the conservation of angular momentum is applied for a certain gas mass, then $v_1 r_1 = v_2 r_2$ (for m). When r is reduced the speed is increased. As r is reduced by half, the speed should be doubled. There is no need of rough work for this question. You can do it from your memory. When the distance is reduced from 80 km to 40 km, the speed should increase from 150 km h^{-1} to 300 km h^{-1} . The speed of the hurricane should be increased when it goes to the middle. We know that from the general knowledge. The middle of the hurricane is known as the eye.

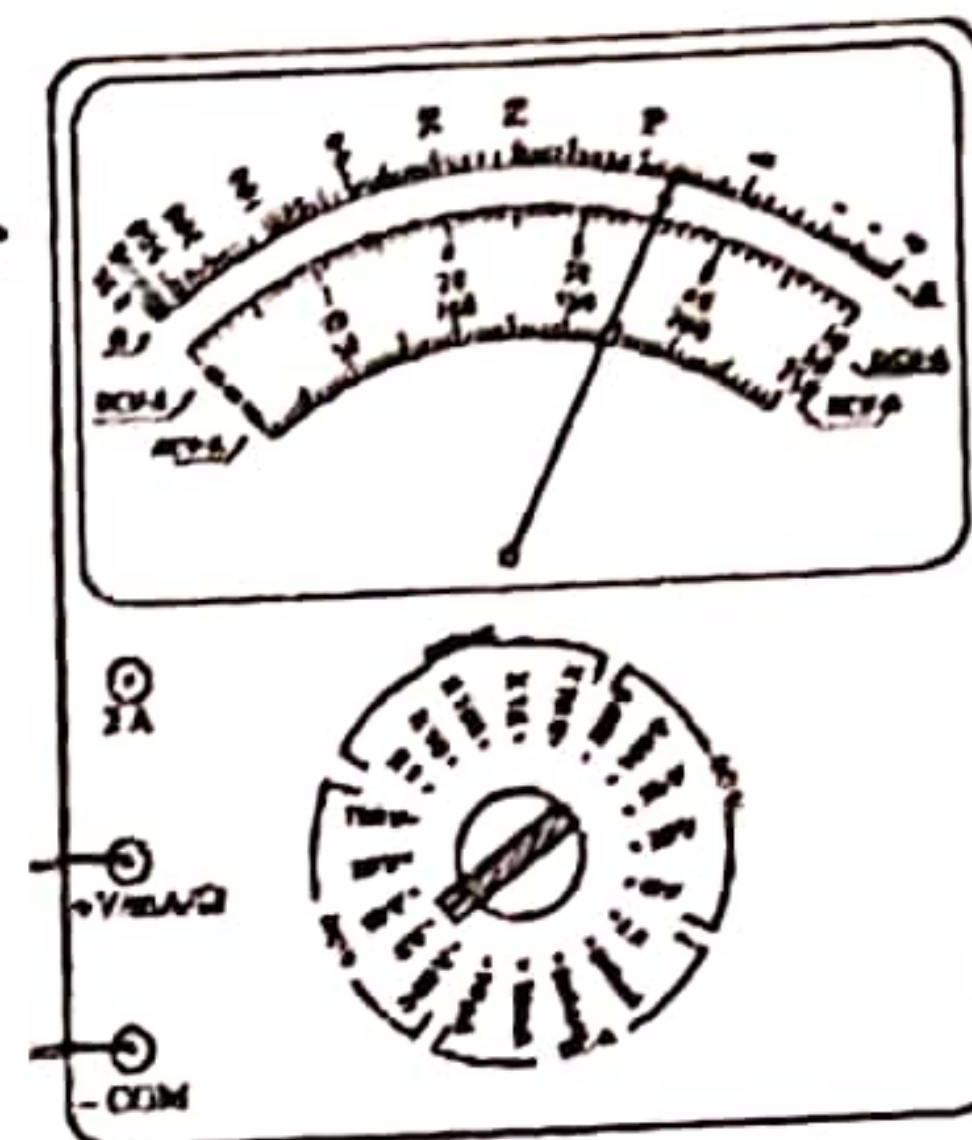
When a hurricane is blown, gradually the speed of the air is increased. When it is passing the eye region, it gets calm down and again the wind is blown to the other side. Actually, the eye is not a single point. According to the figure, it is normally a circular area with a diameter of (30 – 65) km. This place is a very calm place. The boundary that it is covered is known as the eye wall. A higher wind speed is there at this boundary.



When we go across the eye region, we think that the hurricane is over. But it is not like that. Next, it tends to blow winds to the other direction with a high speed. Even if we feel the calmness, then the next moment it gets violent.

26. An analogue multimeter connected to a circuit is shown in the figure. The reading of the multimeter is

- (1) 8 Ω (2) 7 mA (3) 1-4 V
(4) 7 V (5) 14 V



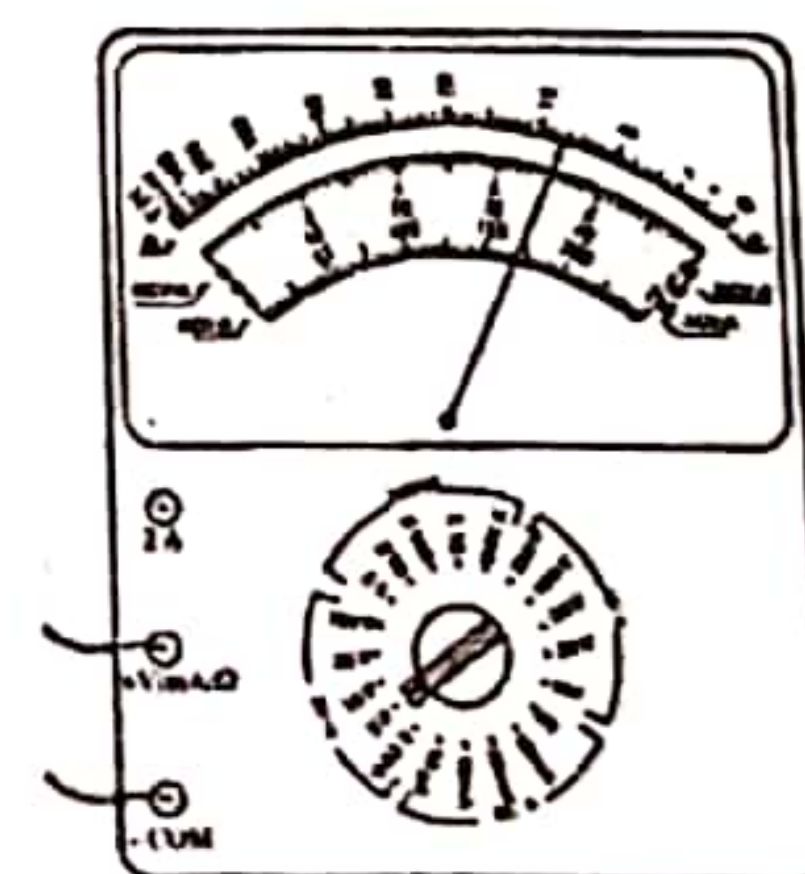
Moving Coil Meters

08

The figure has shown an analog multimeter. A multimeter is a meter that can measure direct voltages and currents (DCV-A) as well as alternative voltages and currents (ACV-A) along with resistor values (Ω). By rotating the lower knob to the required place of the multimeter, you can select the type of measurement you want and the limit. There is an arrow or a small dug hole in the adjusting knob which we rotate. In some multimeters the adjusting corner has made into a triangle.



The multimeter that is shown here has set the knob to 2 V DC-V. It indicates that it measures a simple voltage and the limit it has set up for 2V range. Now look at the DC-V scale. The numbers that are marked there are 0-10 V or 0- 50 V. According to 10 V reading, the indicator is at 7 V whereas according to 50 V reading the indicator is at 35 V reading. The true reading is not any of the above. Why? The lower knob is set to the complete scale of 2V. Therefore, if the reading is 7 V, then the real measurement is $(2/10) \times 7 = 1.4$ V.



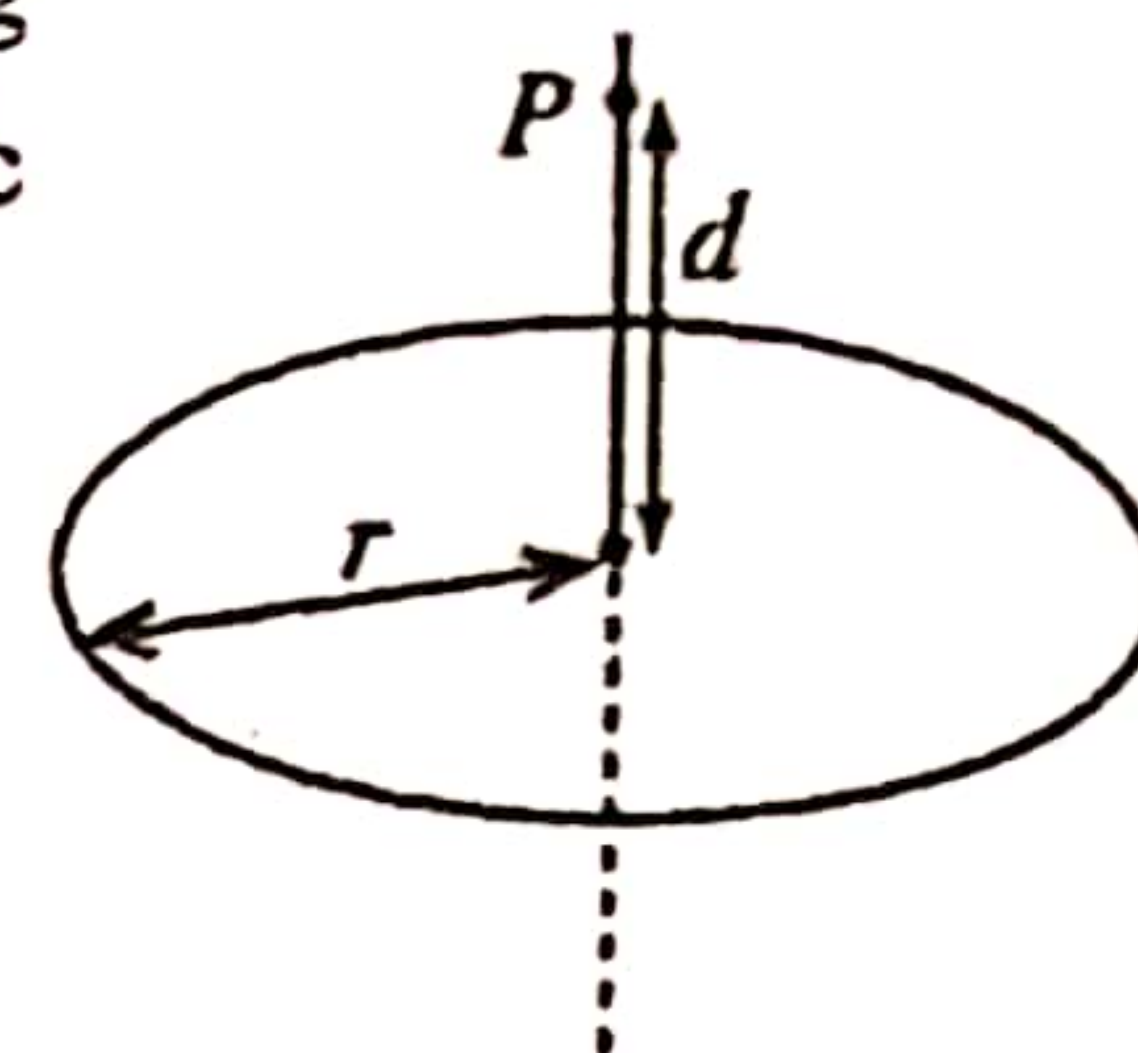
You need to consider here that if 10 V is equivalent to 2 V, then what should be the reading for 7 V? You can check for the scale of 50 V too. $(2/50) \times 35 = 1.4$ V.

Most of the time, children might have thought that 7 V as the correct answer. 35 V is not in the answers. If you just think that the total scale is adjusted to 2 V, then the only answer that is lesser than 2V is 1.4 V.

As the real voltage measurement is 1.4 V, a good reading with a greater deflection can be obtained when the scale is set to 2V. If we take the range of 10 V, then the indicator which shows the deflection stops at 1.4. Look at the figure. Then you will get a smaller deflection and the percentage error in the reading will get greater. If it was set to 0.02 V range, then 1.4 V cannot be read. $1.4 > 0.02$. If so, the indicator goes to the maximum end and starts to dance.

Such multimeters are known as AVO meters. This is a technical word (A – Ampere, V -Volt and O- Ohm).

27. A large number of point charges are uniformly spread out over a non-conducting ring of radius r . If the total charge on the ring is Q , what is the electrostatic potential at point P on the axis of the ring, as shown in the figure?



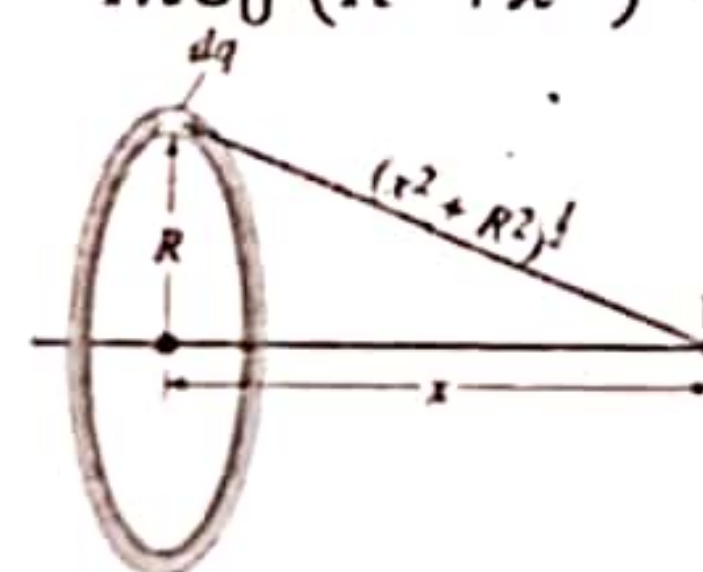
- (1) $\frac{Q}{4\pi\epsilon_0 d}$ (2) $\frac{Q}{4\pi\epsilon_0 r}$
 (3) $\frac{Q}{8\pi^2\epsilon_0 r d}$ (4) $\frac{Q}{4\pi\epsilon_0 \sqrt{r^2 + d^2}}$
 (5) $\frac{rQ}{4\pi\epsilon_0 d \sqrt{r^2 + d^2}}$

02 Electrostatics Potential

This has been given as the 26th question in paper 2006. Finding the potential is easy. The potential is not a vector. We will consider a small dq charge in the ring. The potential that it creates at point

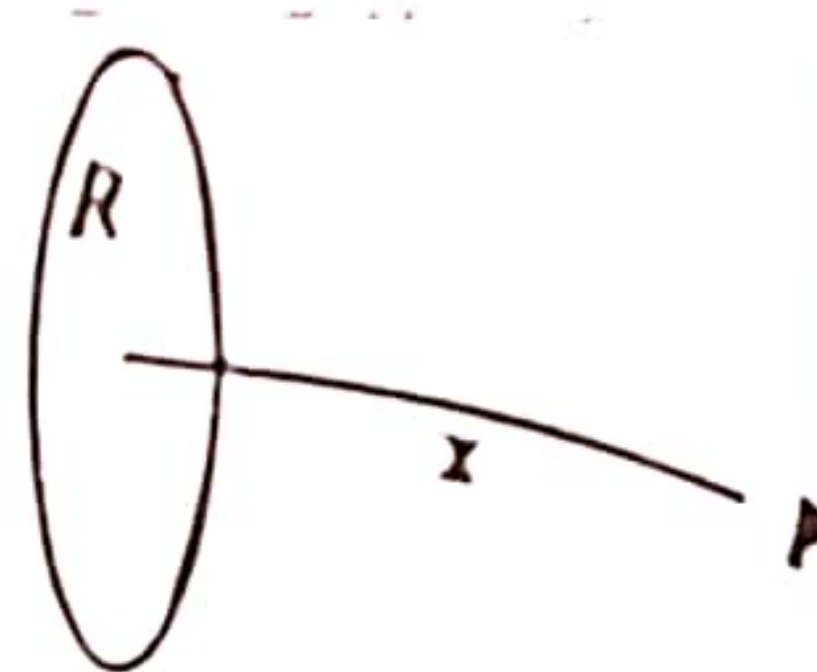
$$P, dv = \frac{1}{4\pi\epsilon_0} \frac{dq}{(R^2 + x^2)^{1/2}}$$

There is same distance from every point charge (small section) to point P . Therefore, the potential created in point P from the charge at the whole ring $V = \sum dv = \frac{1}{4\pi\epsilon_0} \frac{\sum dq}{(R^2 + x^2)^{1/2}} = \frac{1}{4\pi\epsilon_0} \frac{Q}{(R^2 + x^2)^{1/2}}$



• A thin circular ring which has a uniformly distributed charge Q , the electric potential at point P on the axis of the ring is:

$$V = \frac{1}{4\pi\epsilon_0} \int \frac{dq}{r} = \frac{1}{4\pi\epsilon_0} \frac{1}{(x^2 + R^2)^{1/2}} \int dq = \frac{1}{4\pi\epsilon_0} \frac{Q}{(x^2 + R^2)^{1/2}}$$



When $x = 0$, the potential (at the centre of the ring) should be $\frac{1}{4\pi\epsilon_0} \frac{Q}{R}$.

It happens only in the upper expression. When $x \gg R$, then $V = \frac{1}{4\pi\epsilon_0} \frac{Q}{x}$. As $x \gg R$, then $(x^2 + R^2) \approx x^2$. That means when point P is departed relative to R , it is like the total charge of the ring is concentrated to its centre.

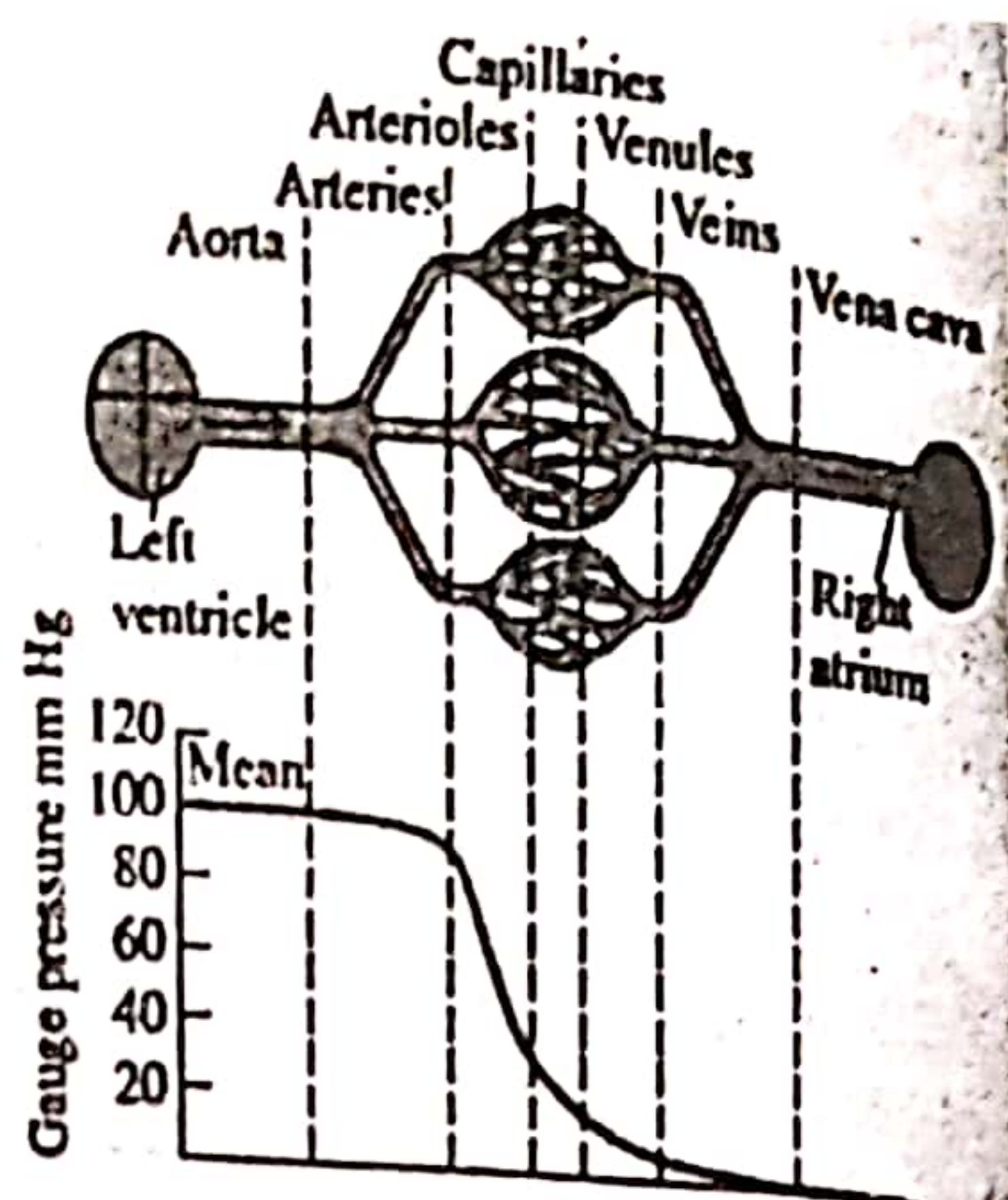
28. The human blood circulatory system consists of about one billion (10^9) capillary vessels each with an average diameter of $8 \mu\text{m}$. If the blood is pumped from the heart at a rate of 5 litres per minute, what is the average speed of blood flowing through the capillary vessels in cm per minute?

- (1) $\frac{1}{32\pi}$ (2) $\frac{25}{16\pi}$ (3) $\frac{25}{4\pi}$ (4) $\frac{125}{16\pi}$ (5) $\frac{125}{4\pi}$

02 Hydrodynamics

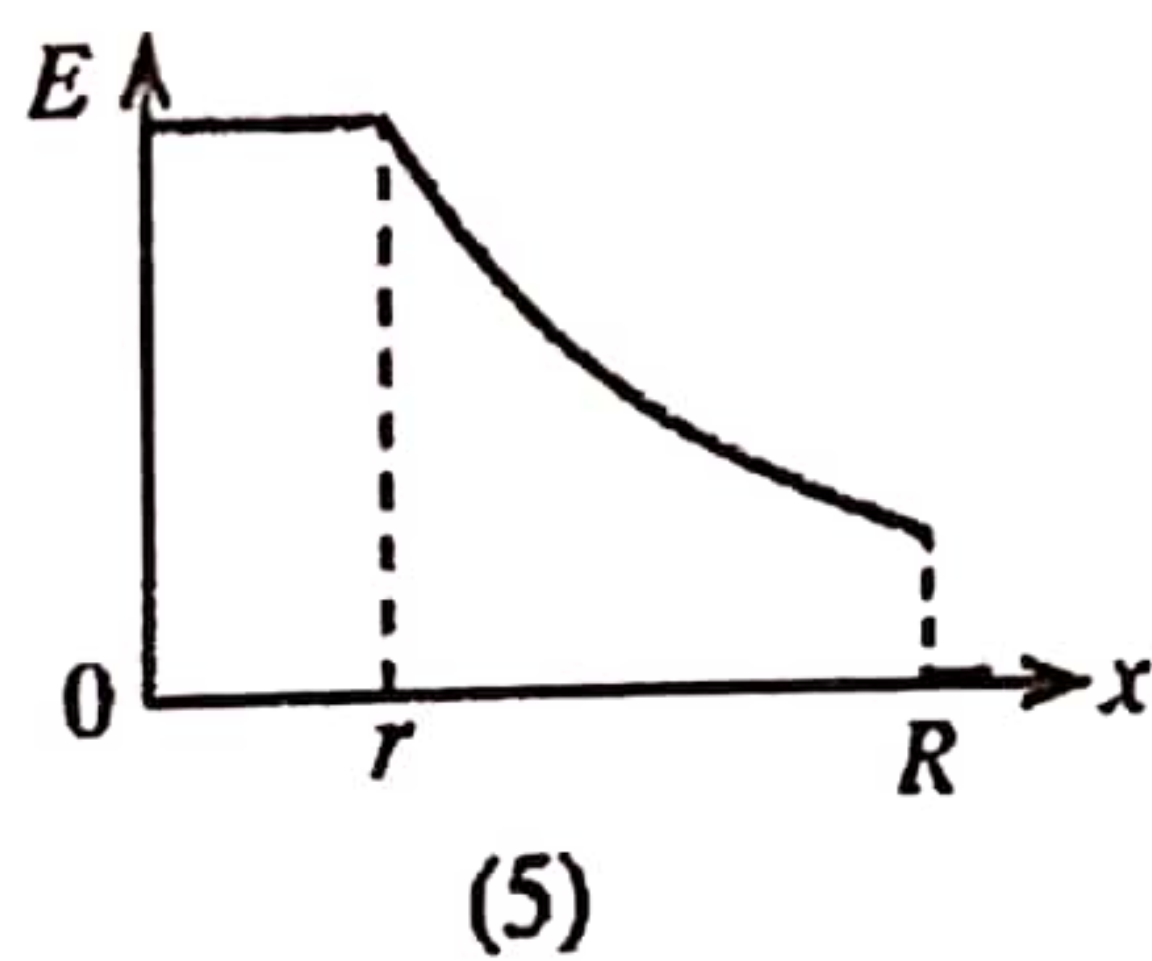
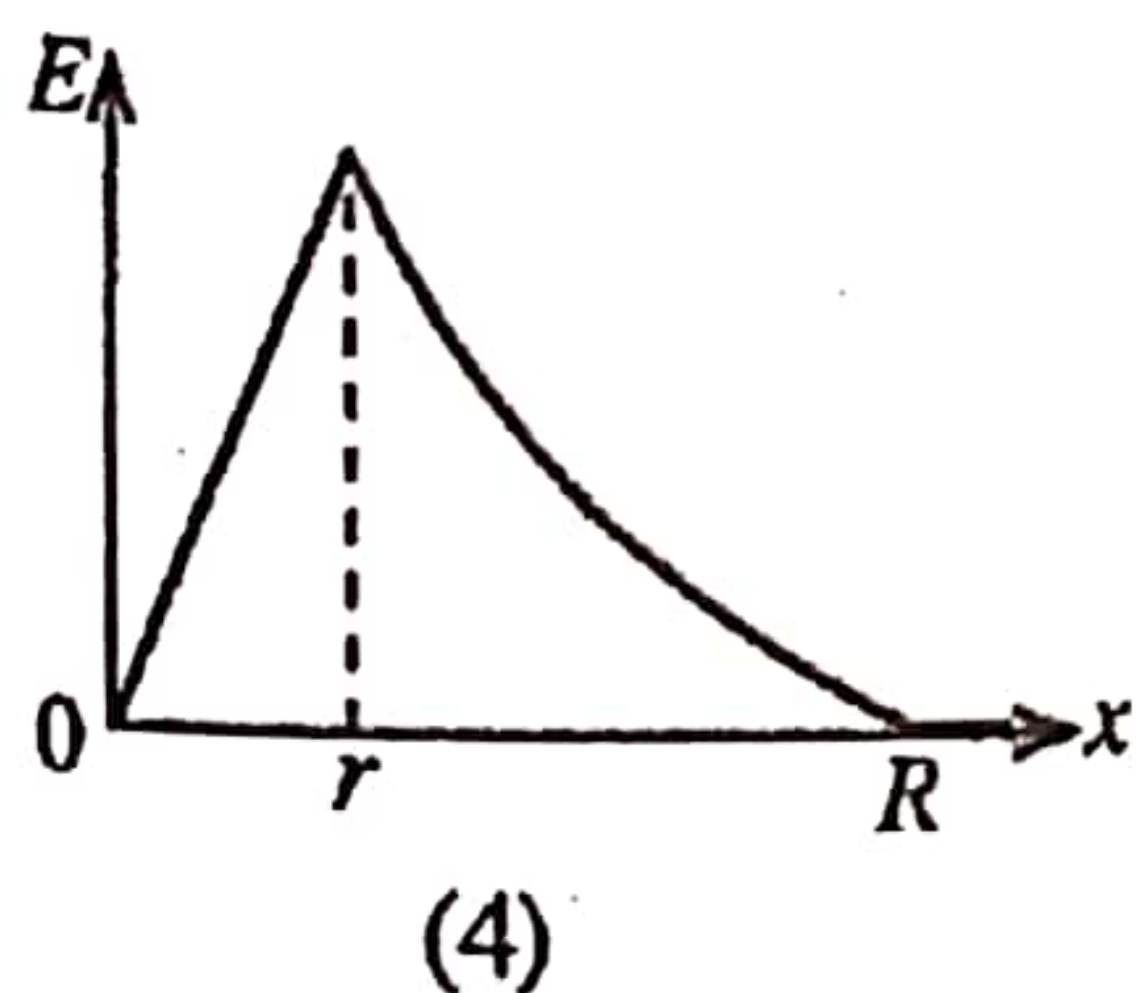
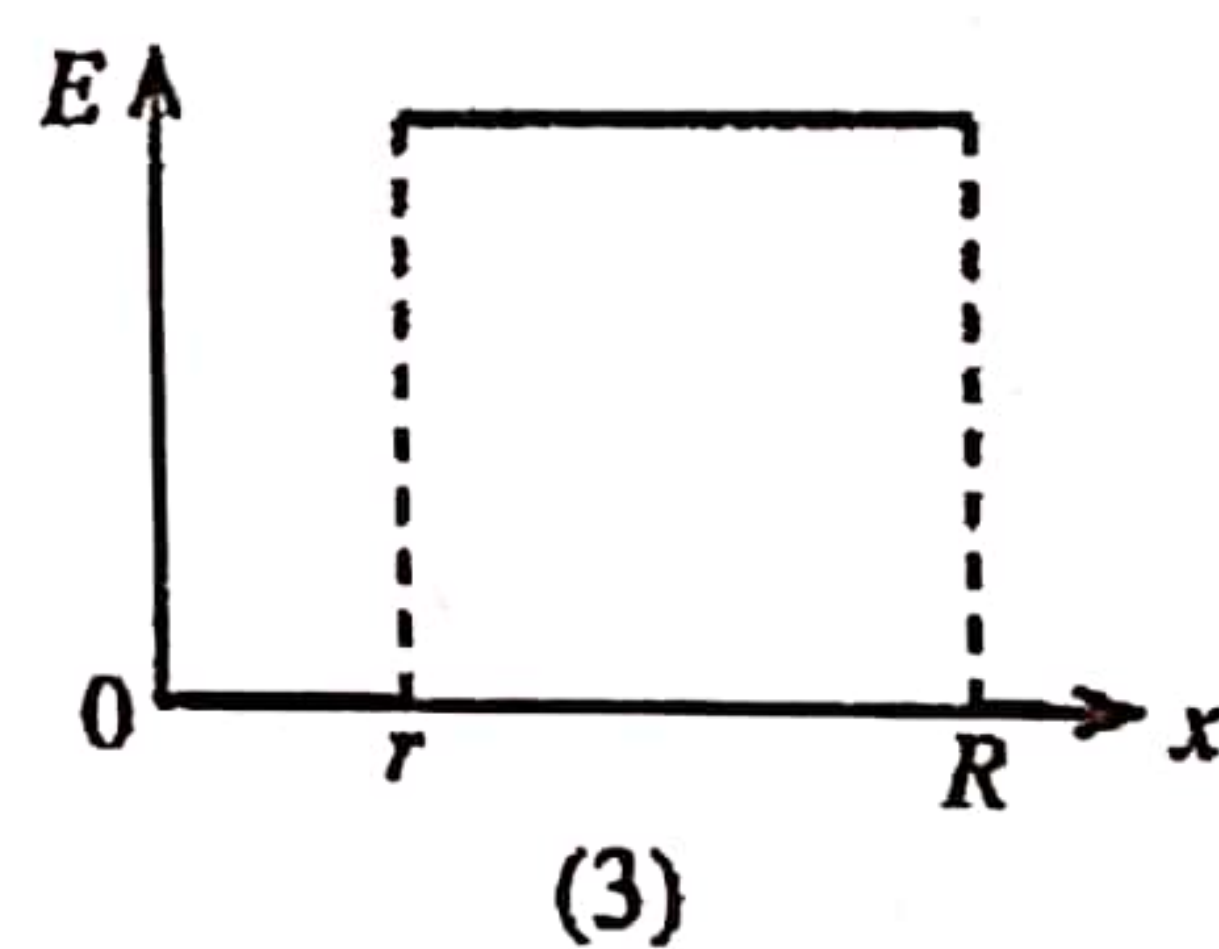
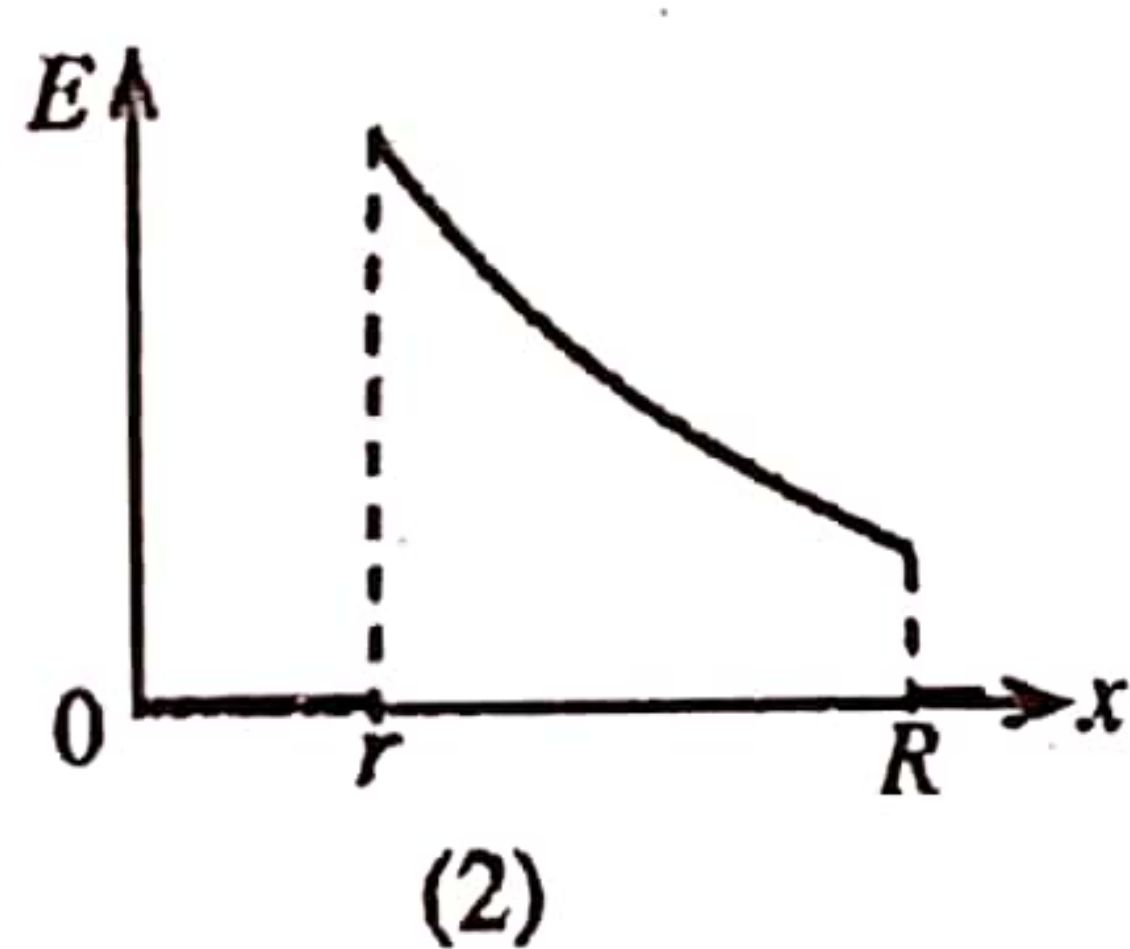
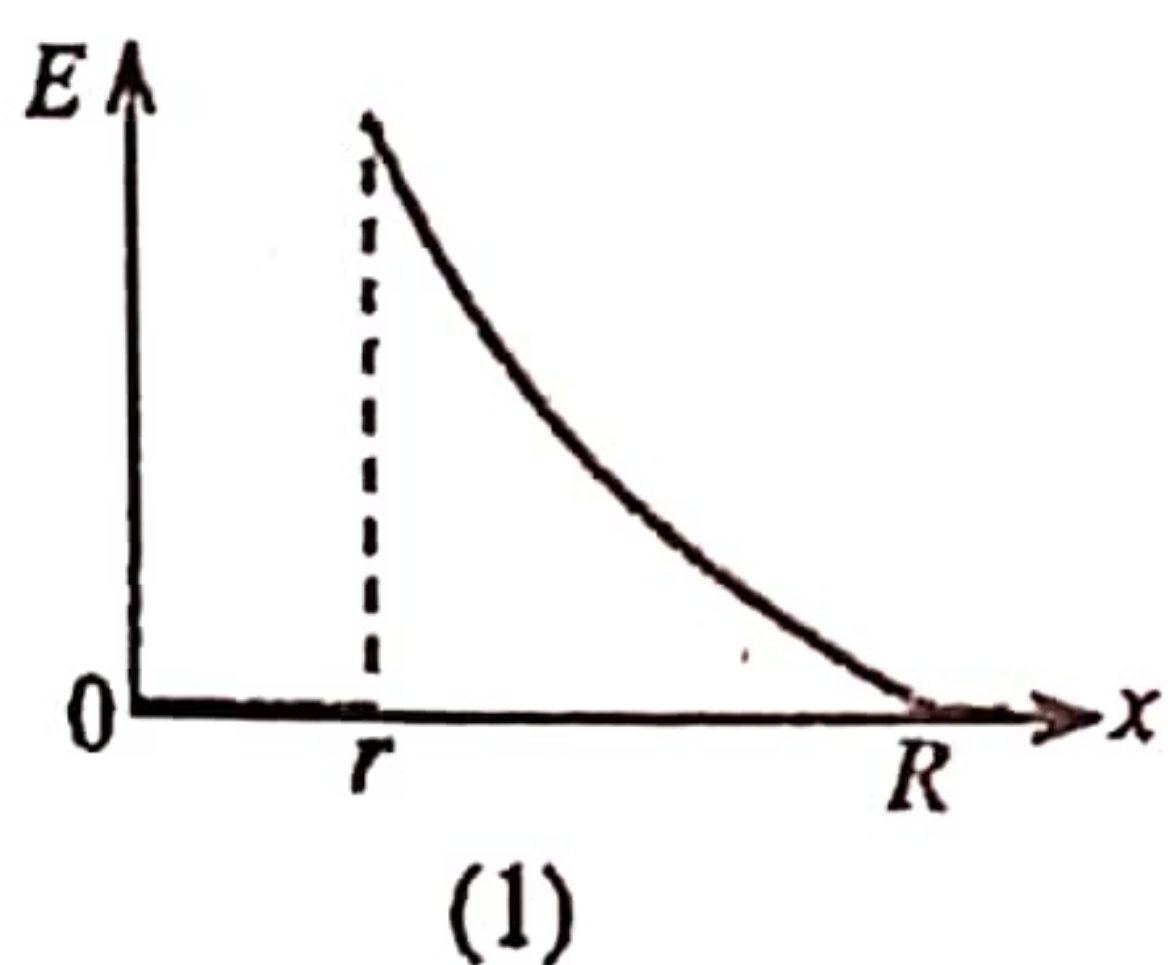
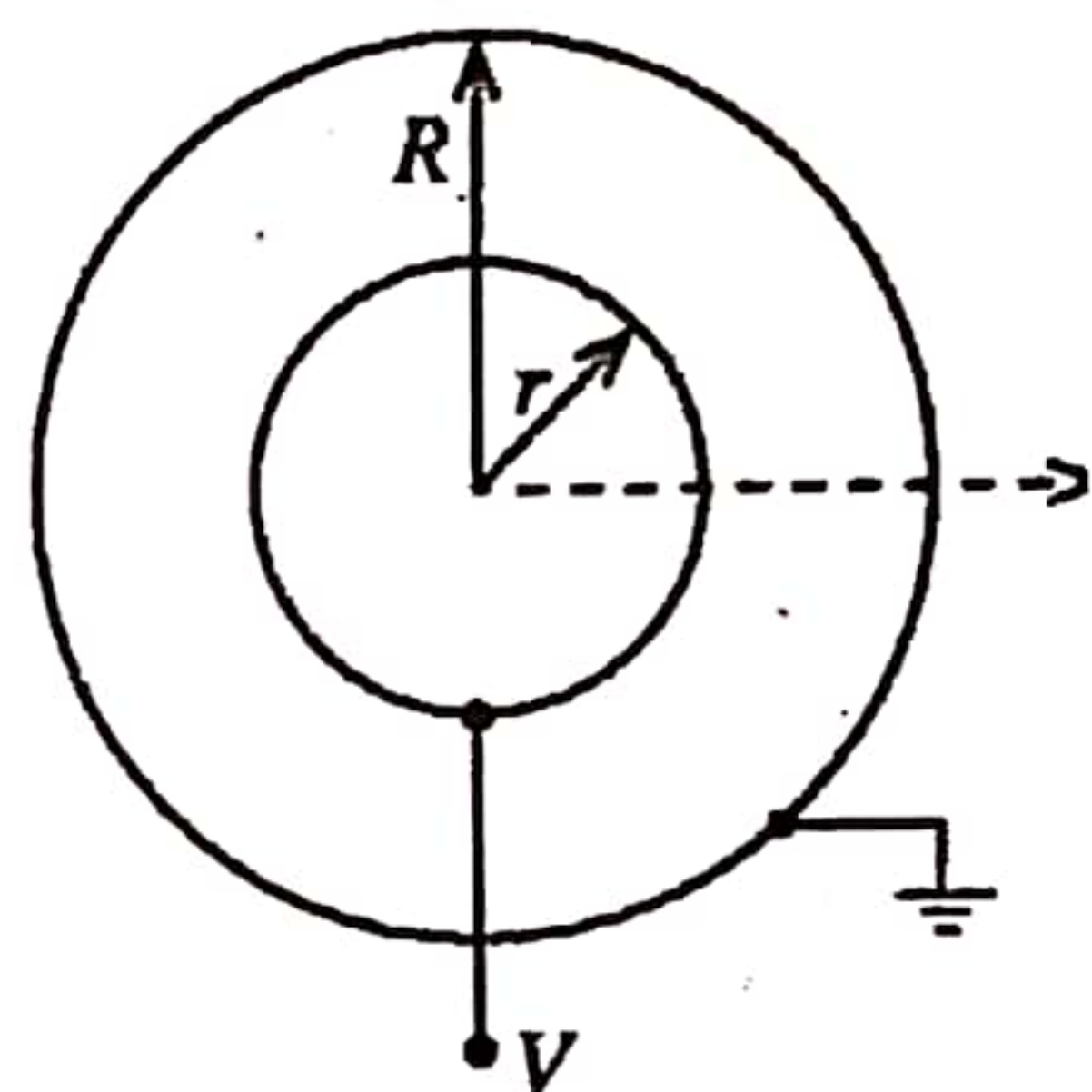
As shown according to the figure, the blood that is flown from aorta are divided into arteries and then goes to arterioles. Then they go across lot of capillaries in the body which then goes to venules, veins and then comes back to heart across vena cava. If 5 l of blood is pumped by the heart per minute, then its rate is 5000 cm^3 per minute ($1 \text{ l} = 10^3 \text{ ml} = 10^3 \text{ cm}^3$). In such problems, it is important that you need to keep in mind that

1 cm^3 is 1 ml. The amount of blood is divided into 10^9 capillaries each with a cross sectional area of πr^2 . If v is the average speed in a capillary cm per minute, then $\pi (4 \times 10^{-4})^2 v \times 10^9 = 5 \times 10^3 \rightarrow v \times \pi \times 16 = 500$
 $\rightarrow v = 500/16\pi = 125/4\pi$





Here l, μm , cm should not be confused with each other. Liters should be converted into cm^3 . As the radius is given in μm , it should be converted to cm ($1 \mu\text{m} = 10^{-6} \text{ m} = 10^{-4} \text{ cm}$). Then v is obtained in cm min^{-1} . Look at the 09th question of paper 2010.

29. Two thin spherical metallic shells are placed concentrically as shown in the figure. Inner shell is kept at a potential U while the outer shell is grounded. The variation of the electric field E with distance x from the centre is best represented by



Gauss Theorem

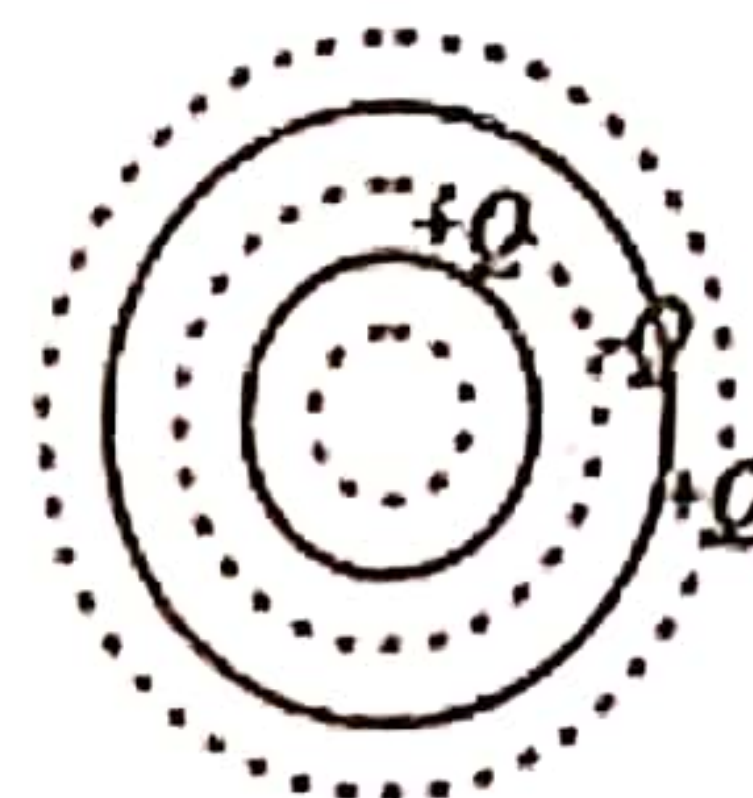
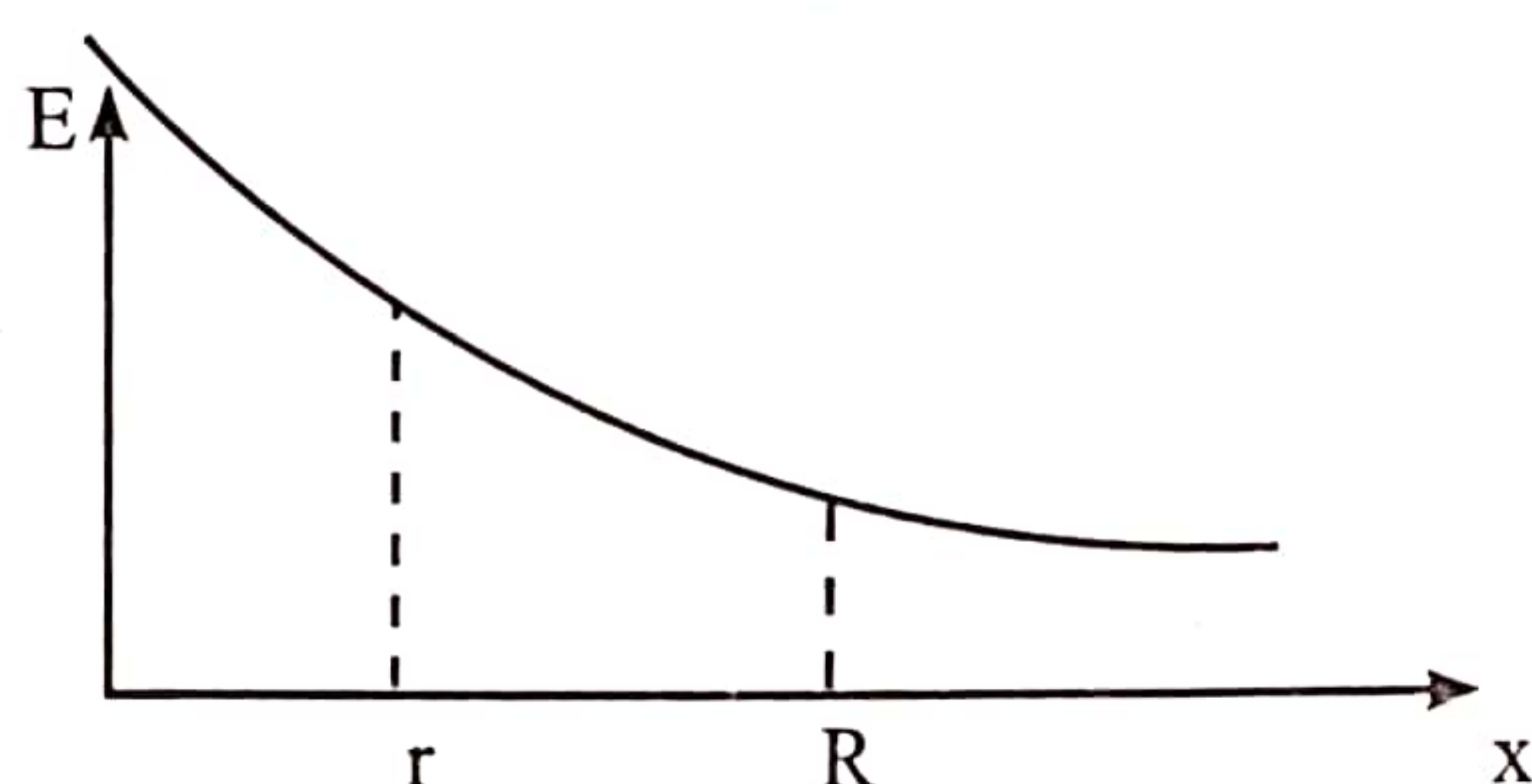
06

To obtain negative charge	To obtain positive charge
	
Conductor earthed by touching it with finger.	Conductor earthed by touching it with finger.
➤ Closed path for <u>electrons to flow from Earth to neutralise positive charge at Q.</u>	➤ Closed path for <u>electrons to flow from conductor to Earth.</u>

Look at the 58th question of paper 1999. This is the same question.

Keeping the inner shell at $+V$ potential means that there is a $+$ charge ($+Q$) in its surface. If so, then the inner surface of the outer shell gets $-Q$ whereas the outer surface gets $+Q$ by induction. We know that E is zero inside the inner shell.

If we consider a Gaussian surface in the inner shell, then there is no net charge inside that surface. Therefore, $E = 0$. E in between the shells varies according to $1/r^2$. The net charge inside the Gaussian surface outside the shells $= +Q - Q + Q = +Q$. Therefore, if we consider that there is no thickness in the outer shell, then the variation of E for the above arrangement is like this.

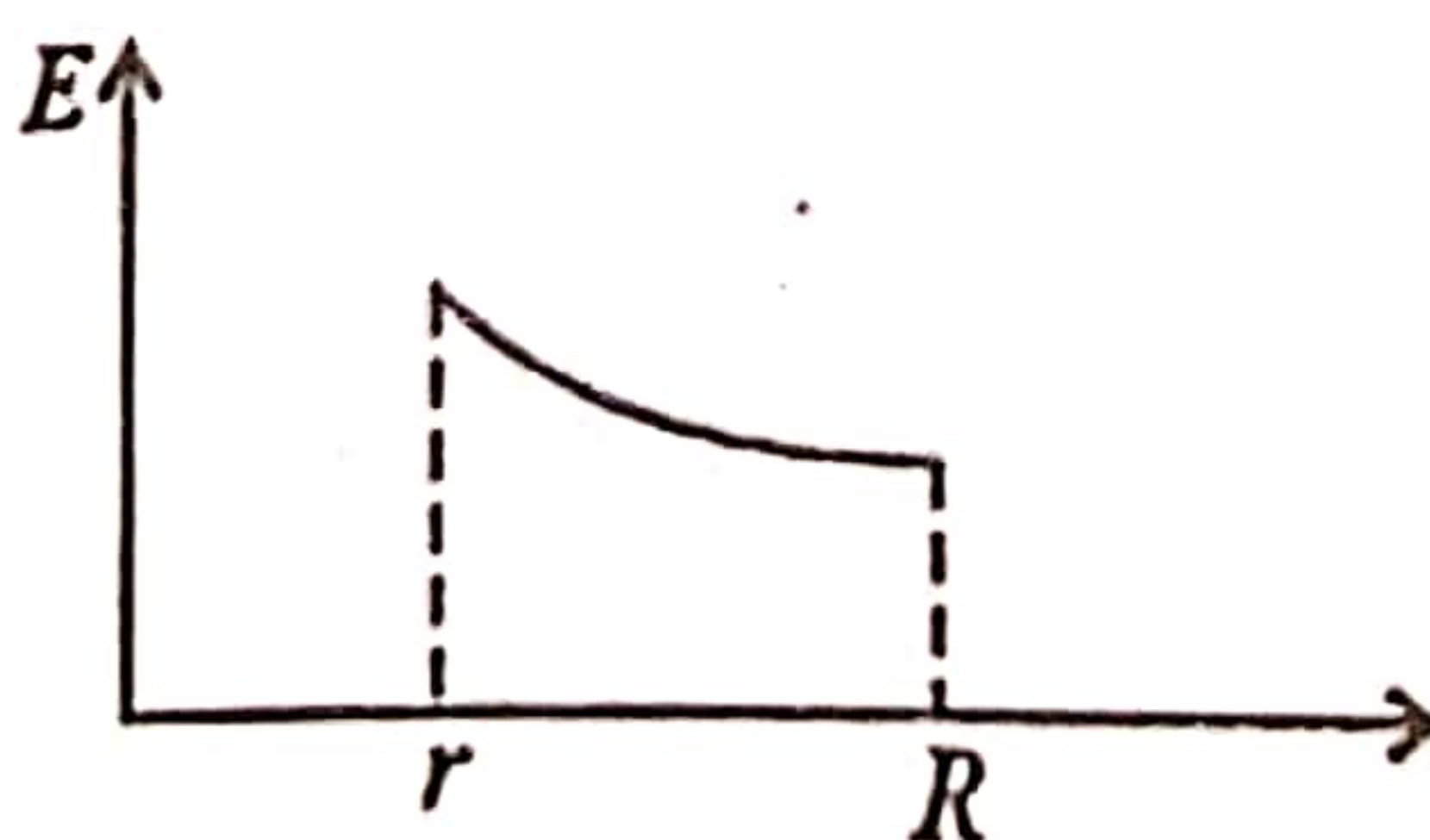


Now if the outer shell is earthed, then negative charges will flow from the earth and cancel off the $+Q$ charge which is outside the shell. Even we say that the positive charge goes to earth, what really happens is that the electrons come from the earth and cancel off the positive charge. The earth is considered as a source with more electrons or as a sink which can remove extra electrons. Practically its electric potential does not change by taking electrons or giving electrons to earth. Does it due to this reason why the earth is known as 'Mother Earth'? Are not our mothers like this? Our mothers can cope anything as well as love us unconditionally.

On the other hand, the shells are at positive potential (before earthing). The earth is at zero potential. Therefore, the electrons should flow from zero potential to positive potential.

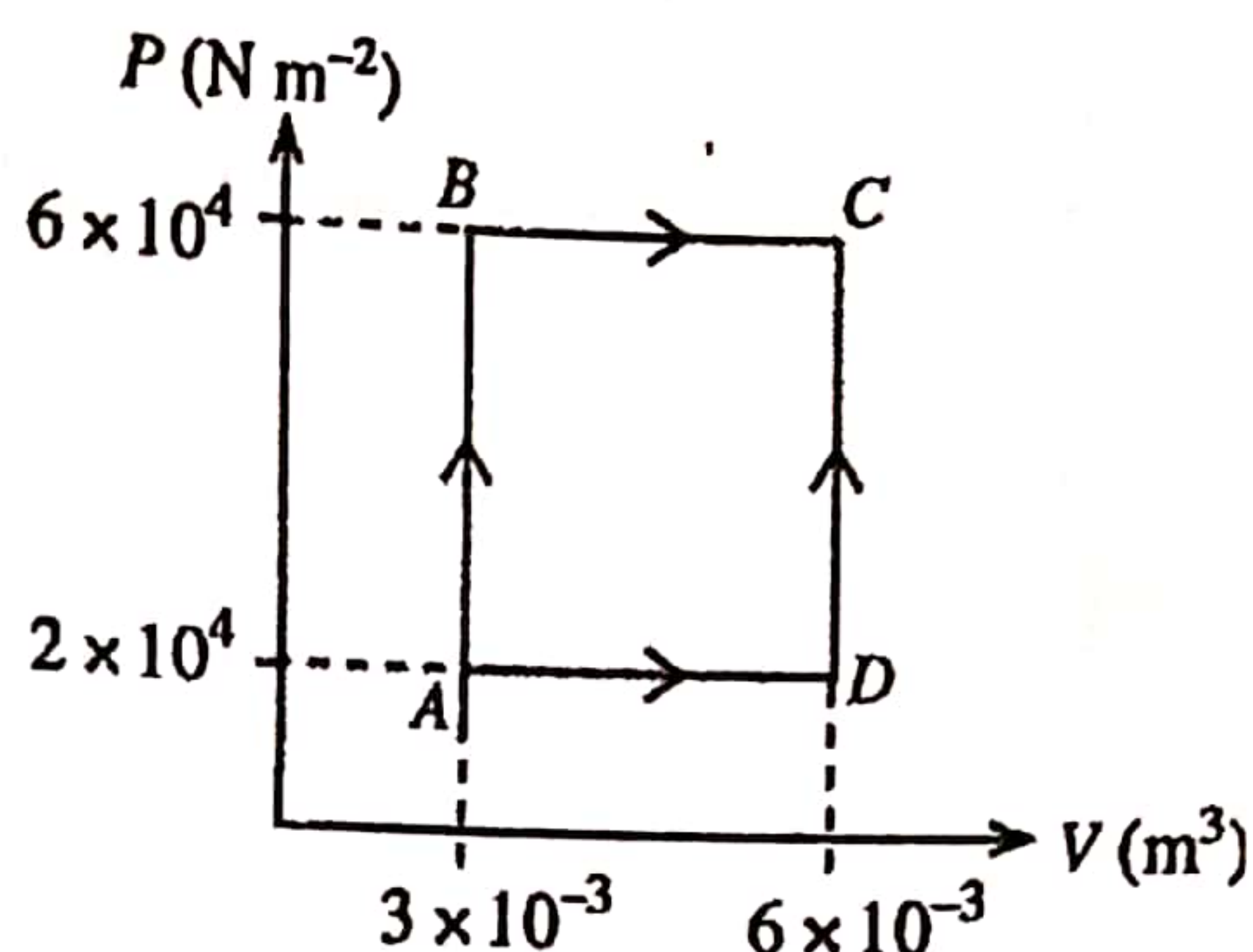
Now when the outer shell is earthed, the outside electric field intensity gets suddenly zero. Now the net charge inside the Gaussian surface which is drawn outside gets zero ($+Q - Q = 0$). Therefore, the variation of E like this way.

At $x = R$, suddenly it drops to zero. It is not a journey which was intended to drop to zero. It is felt that E should be zero when it comes to the outer shell. In such an occasion, how can the electric potential V vary?



30. An ideal gas expands from state A to state C along two different paths, ABC and ADC, as shown in the P-V diagram. The heat absorbed by the gas during the processes AB and BC are 200 J and 700 J, respectively. What is the change in internal energy, when the gas expands along the path ADC?

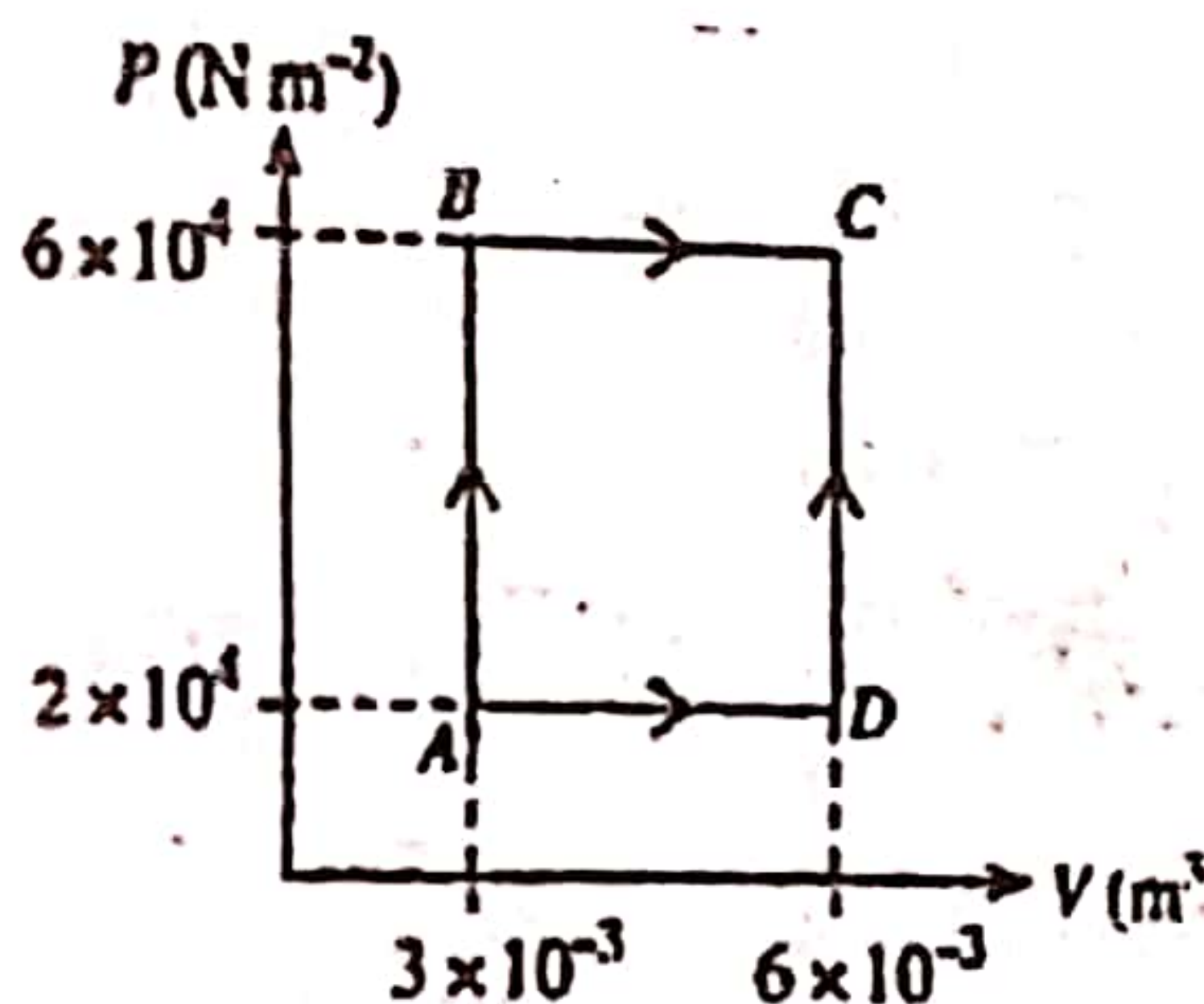
- (1) 380 J (2) 520 J
(3) 720 J (4) 880 J
(5) 1080 J



Thermodynamics

04

Different kinds of P-V curves can be seen in the past papers. The internal energy change (ΔU) is dependent upon the initial and the final state. It does not depend on the path from the initial to the final state. There has been given an absorbed heat for the path of ABC. From that, if ΔU can be found, then even it goes across ADC you get the same ΔU .



From A to B, $\Delta W = 0$ (as the volume is constant). From B to C, $\Delta W = P \Delta V$ (the pressure is constant, it is an isobaric process). If we apply $\Delta U = \Delta Q - \Delta W$ for ABC, then

$\Delta U = 200 + 700 - [6 \times 10^4 (6 - 3) \times 10^{-3}] = 900 - 180 = 720$ J. The heat is being absorbed in both occasions. Therefore, ΔQ is positive whereas ΔW is also positive. As the volume is increased, ΔW is positive.

31. A ball is dropped freely to a floor from a height of 1 m. If its speed is reduced by 25% at each bounce, what would be the height the ball reaches after three bounces?

- (1) $\frac{3}{4}$ m (2) $\left(\frac{3}{4}\right)^2$ m (3) $\left(\frac{3}{4}\right)^3$ m (4) $\left(\frac{3}{4}\right)^6$ m (5) $\left(\frac{3}{4}\right)^9$ m

Linear Motion

02

Do not try to solve this in a lengthy way. If v is the speed before it hits the ground, then $mgh = \frac{1}{2}mv^2$. So, h is proportional to v^2 . $h \propto v^2$

In a bounce back, if the speed is reduced by 25%, then what is left will be 75%. That means $\frac{3}{4}$. All the answers are given in the powers of $\frac{3}{4}$. After three times of bounce back, the speed is $\left(\frac{3}{4}\right)^3$. But as $h \propto v^2$, h goes with $\left(\frac{3}{4}\right)^6$. Initially as $h = 1$ m, the height that the ball rises will be $\left(\frac{3}{4}\right)^6$ m.

32. Part of an orbiting satellite is coated with a metal that has a work function of 5 eV. The Planck constant is $4.1 \times 10^{-15} \text{ J s}$ and the speed of light is $3 \times 10^8 \text{ m s}^{-1}$. What could be the longest wavelength of incident sunlight that can eject an electron from the metal coating?

- (1) 123 nm (2) 246 nm (3) 683 nm
(4) 800 nm (5) 1230 nm

11

Photoelectric Effect

It is very easy. Even there was such a question in the model paper. The longest wave length means the threshold wavelength. Then $\phi = hc/\lambda_0 \rightarrow \lambda_0 = (4.1 \times 10^{-15} \times 3 \times 10^8)/5 = 2.46 \times 10^{-7} \text{ m} = 246 \text{ nm}$ ($2.46 \times 10^{-7} \times 10^9 \text{ nm}$)

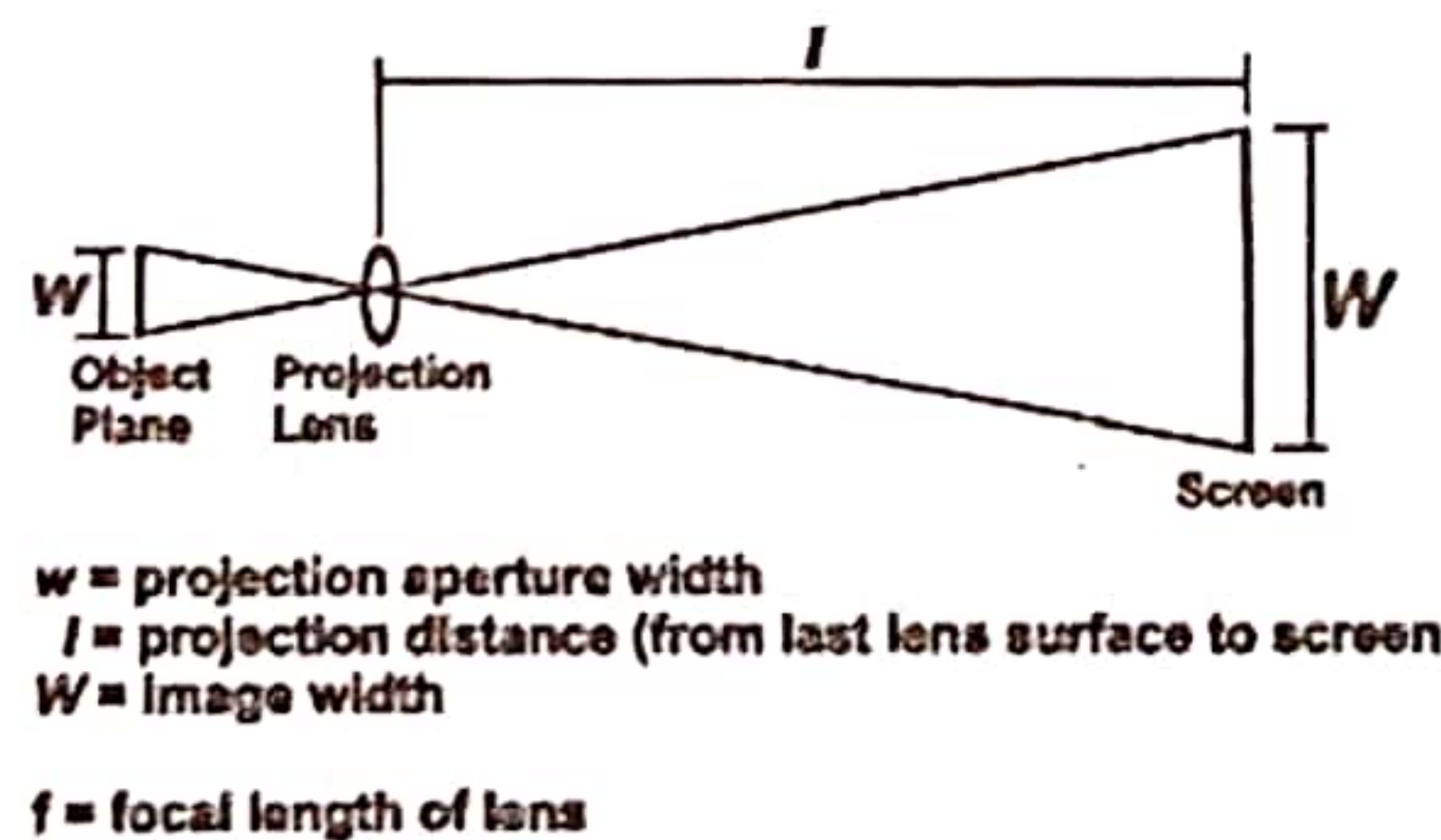
33. A standard photographic slide has a picture size of 30 mm x 40 mm. An enlarged image of the slide is projected onto a screen 4.0 m away from the projection lens of a 'single-lens slide projector'. If the size of the image on the screen is 1.2m x 1.6 m, what should be the focal length of the projection lens?

- (1) 4.9 cm (2) 9.8 cm (3) 10.2 cm
(4) 49 cm (5) 98 cm

03

Refraction Through Lenses

Do not think of the area of the figure. The figure size is 30 mm X 40 mm. The image size is 1.2 m X 1.6 m. These two dimensions have been magnified by 40 ($1.2 \times 10^3/30 = 40$ and $1.6 \times 10^3/40 = 40$) times. Therefore, think only about the linear magnification. According to $v/f = 1+m$, $4/f = 41$; $f = (4/41) \times 100 = 9.76 \text{ cm} \approx 9.8 \text{ cm}$

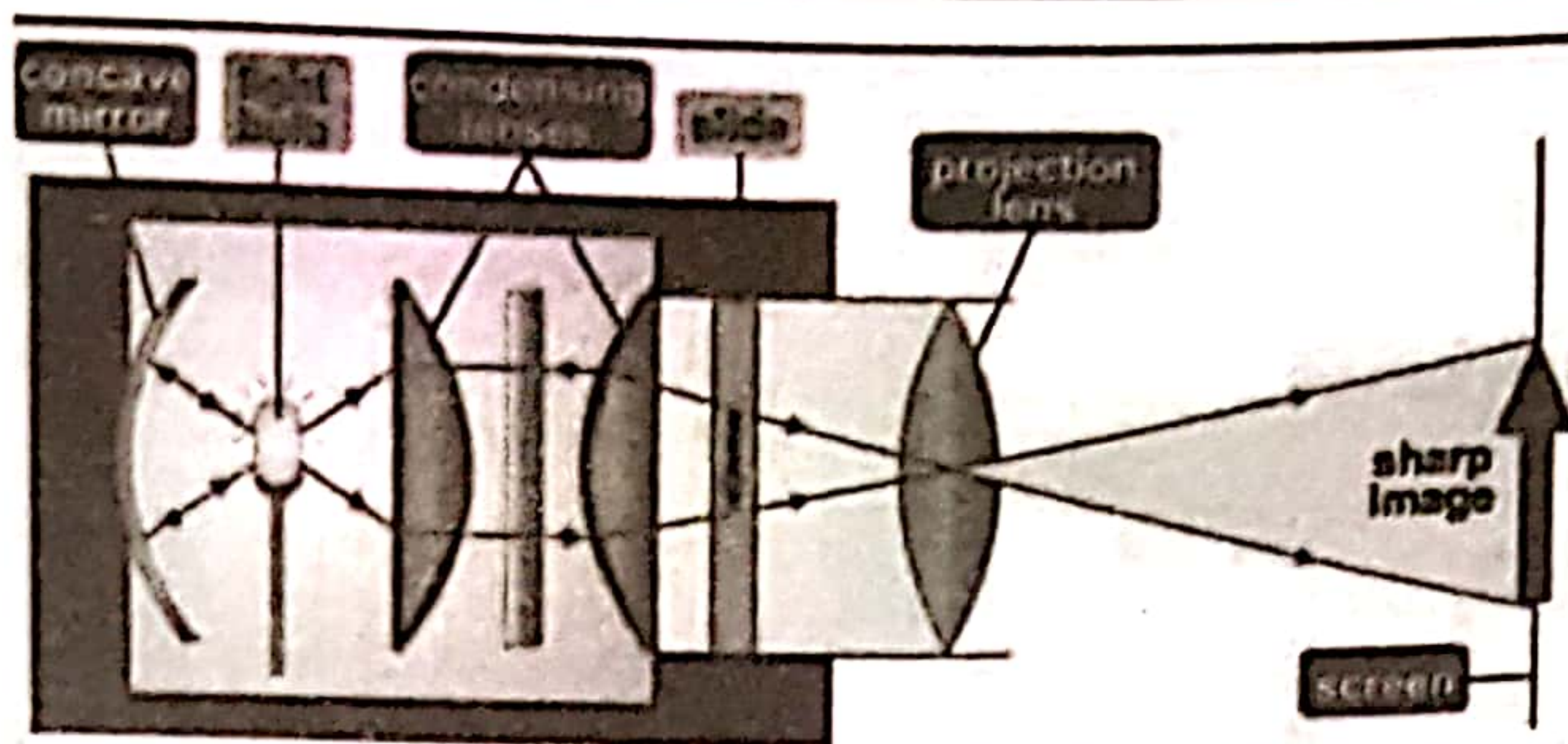


When you see 400/41, it is clear the answer should be near to 10 but less than 10. The only answer which satisfies this is 9.8 cm.

[If we apply lens formula $1/V - 1/U = 1/f$ and sign convention, then $-1/V - 1/U = -1/f \rightarrow 1/V + 1/U = 1/f \rightarrow 1 + V/U = V/f \rightarrow 1 + m = V/f$]

If you think of the area of the figure, then the area has become increased by $40 \times 40 = 1600$ times. We do not do calculations using area magnification. However when the same object is kept at a distance, the height also should be magnified along the magnification of the width in a rectangular shape. There cannot be two types of magnifications.

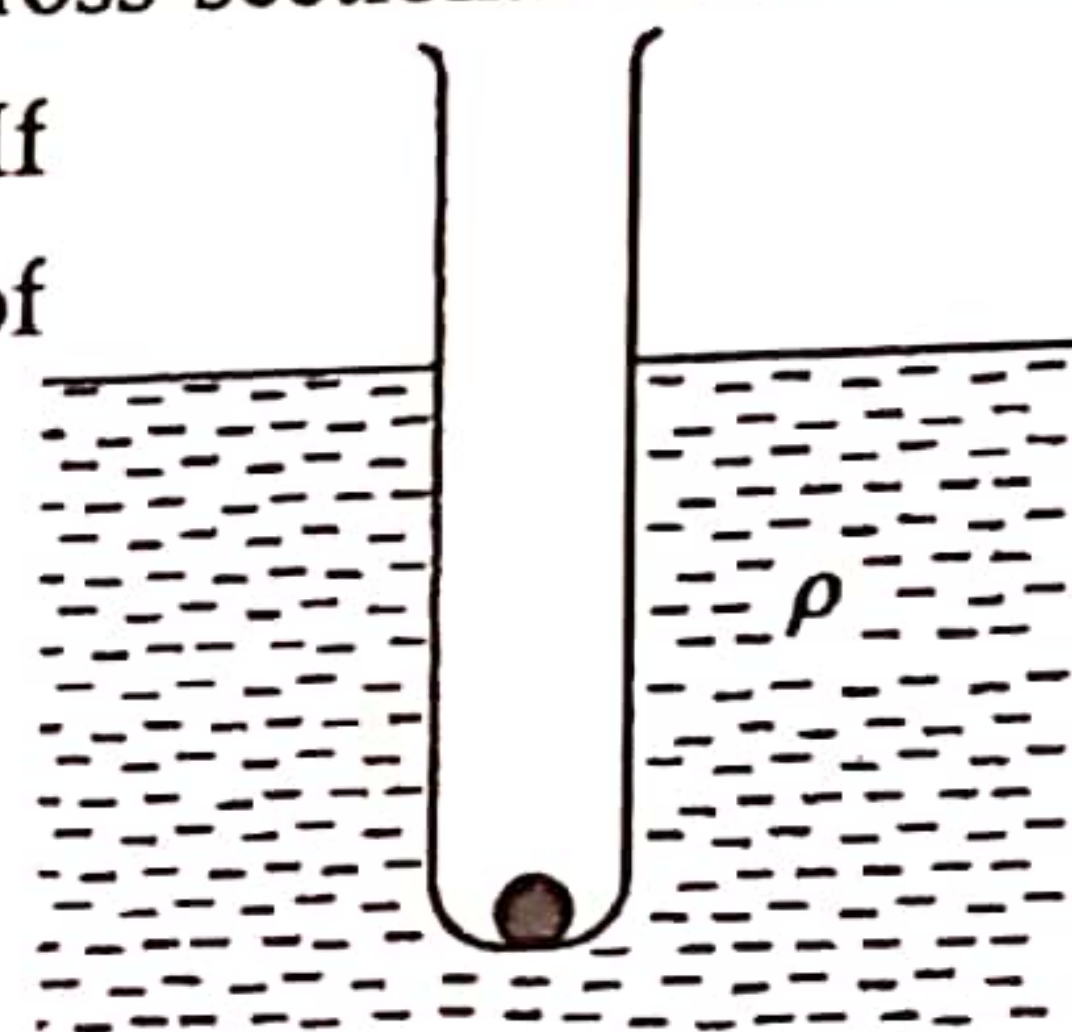
Another important fact is that, when the object is projected on the screen, the object and the image are on placed beside the lens. Therefore, V is negative. You cannot use $m = 1 + d/f$ as in a simple microscope. In a simple microscope you will get a magnified unreal image. Therefore, V is positive. But in a slide projector, you need a magnified real image. If it is not real, then it cannot be projected to a screen. Therefore, if you use $m = 1 + V/f$ by mistake, then you will get $f \approx 10.2 \text{ cm}$.



This figure has shown the ray diagram of a slide projector that is being normally used. The slide should be inserted into the projector by upside down. Then you will get a non-inverted upright image of the object. By using concave mirror and condensing lenses the light is allowed to be incident on the slide with a possible intensity. The bulb is kept at the centre of curvature of the concave mirror as well as in the focus of the first plane convex lens. You will understand that why it has been kept like that way.

34. A test tube is made to float upright in a fluid by placing a metal ball at the bottom of the tube as shown in the figure. Total mass of the tube and the ball is m , density of the fluid is ρ , and the cross-sectional area of the tube is A . Effect of surface tension and the viscosity of the fluid can be neglected. If the tube is given a small vertical displacement, what is the period of oscillations of the subsequent motion of the tube?

- (1) $2\pi\sqrt{\frac{A\rho g}{m}}$ (2) $2\pi\sqrt{\frac{m}{A\rho g}}$ (3) $2\pi\sqrt{\frac{2m}{A\rho g}}$
 (4) $2\pi\sqrt{\frac{m}{2A\rho g}}$ (5) $2\pi\sqrt{\frac{mg}{A^2\rho}}$



Simple Harmonic Motion

03

Consider a tube with a weight according to the figure. To find the oscillatory period of the tube's motion when a small vertical displacement is given to the tube, avoid writing equations when the tube is at equilibrium. As shown in the second figure, if the tube is pulled down a distance of x , then the extra upthrust on the tube is $Ax\rho g$. If we consider this extra force and apply $F = ma$ downwards, then $-Ax\rho g = ma \rightarrow a = -A\rho g/m \cdot x$. This equation is like $a = -\omega^2 x$. This represents a simple harmonic motion. $\omega = \sqrt{\frac{A\rho g}{m}}$. Therefore, oscillatory period $T = 2\pi/\omega = 2\pi\sqrt{\frac{m}{A\rho g}}$.

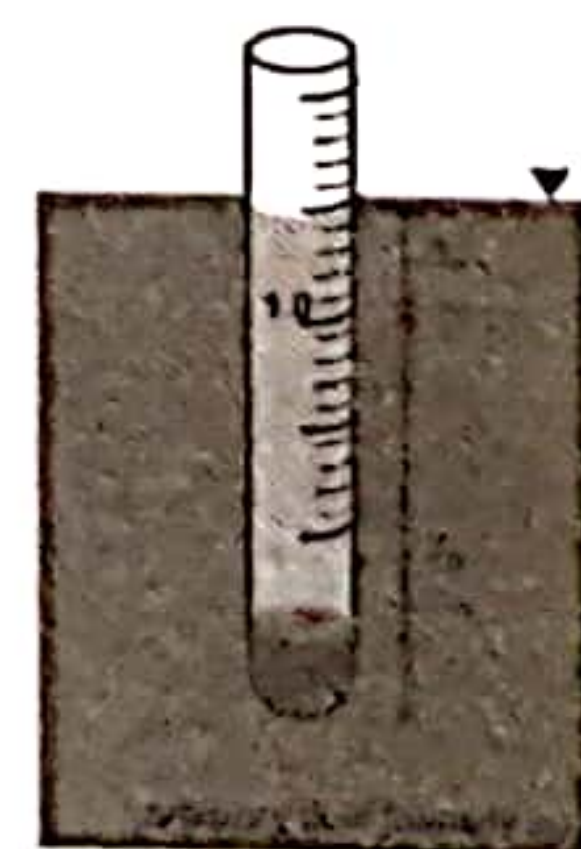
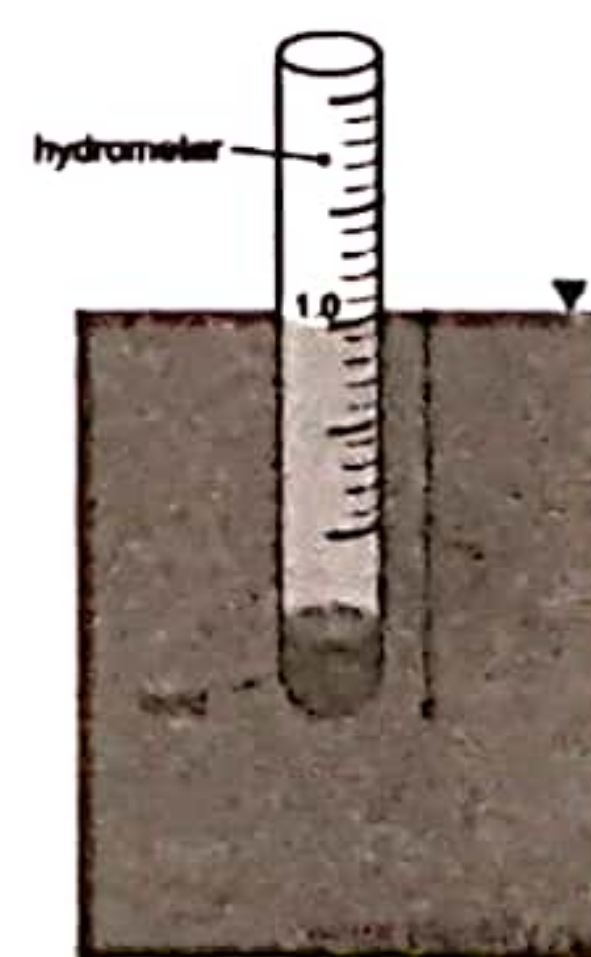
Therefore, oscillatory period $T = 2\pi/\omega = 2\pi\sqrt{\frac{m}{A\rho g}}$. The distance x is measured downwards from the equilibrium level. $\downarrow x$

The extra upthrust is acting vertically upwards. Therefore, when applying $F = ma$ downwards, you need to apply negative sign for F . This extra force is a restoring force. It indicates that the force tries to come into the place where it was before. Do not try to write equations for both occasions. It will consume time. Look at this way. For the equilibrium, $mg = Ax_0\rho g$ (1)

When it is sunk by x distance apply $F = ma$ downwards, $mg - A(x_0 + x)\rho g = ma$ (2) (weight- upthrust)

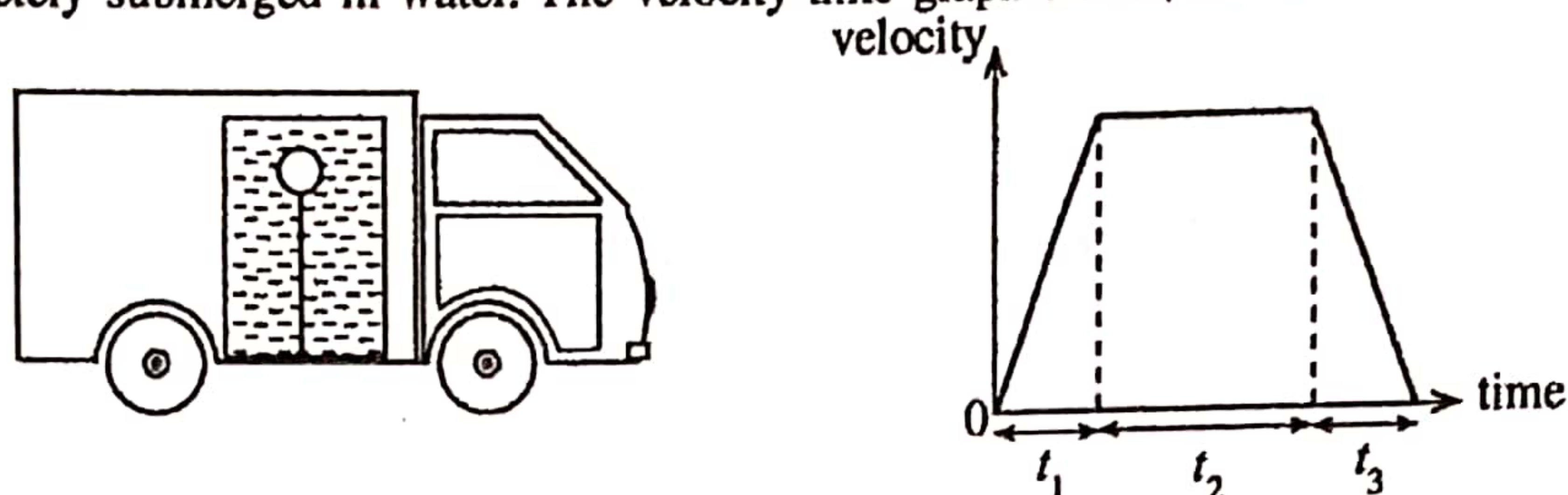
Substitute mg of (2) from (1) $Ax_0\rho g - A(x_0 + x)\rho g = ma$

So, if you directly write equation for the extra force the time can be saved.

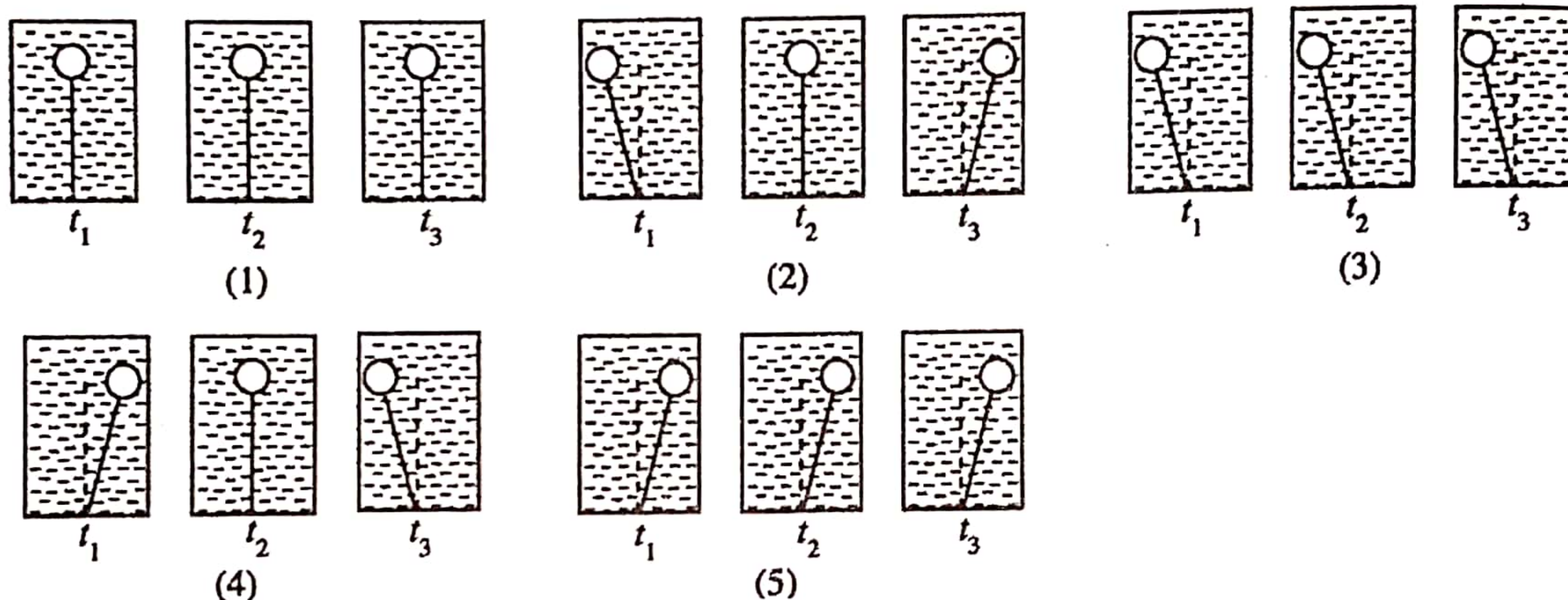


35.

Consider a massless balloon attached to one end of a light string. The other end of the string is attached to the bottom of a water tank which is fixed in a truck as shown in the figure. The balloon is completely submerged in water. The velocity-time graph shows the motion of the truck.



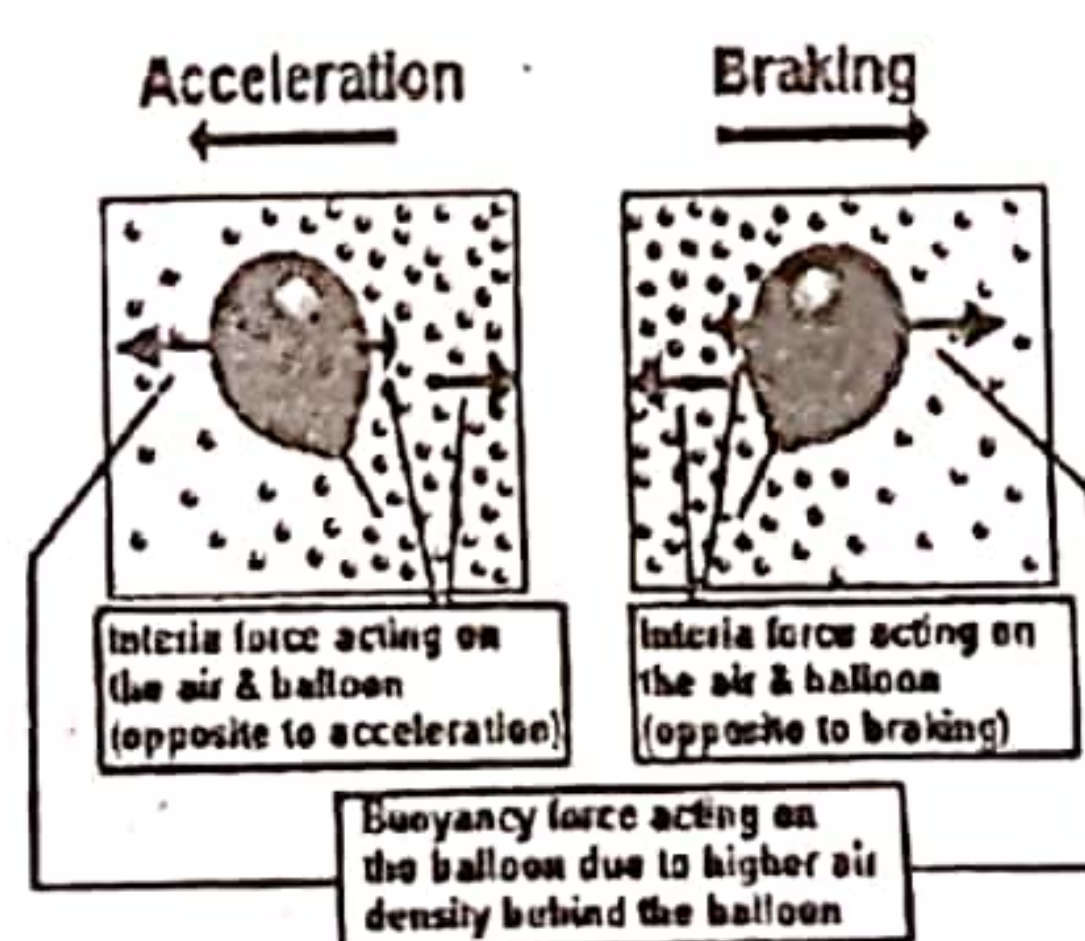
The positions of the balloon and the string inside the water tank during the time intervals t_1 , t_2 , and t_3 are best represented by



02

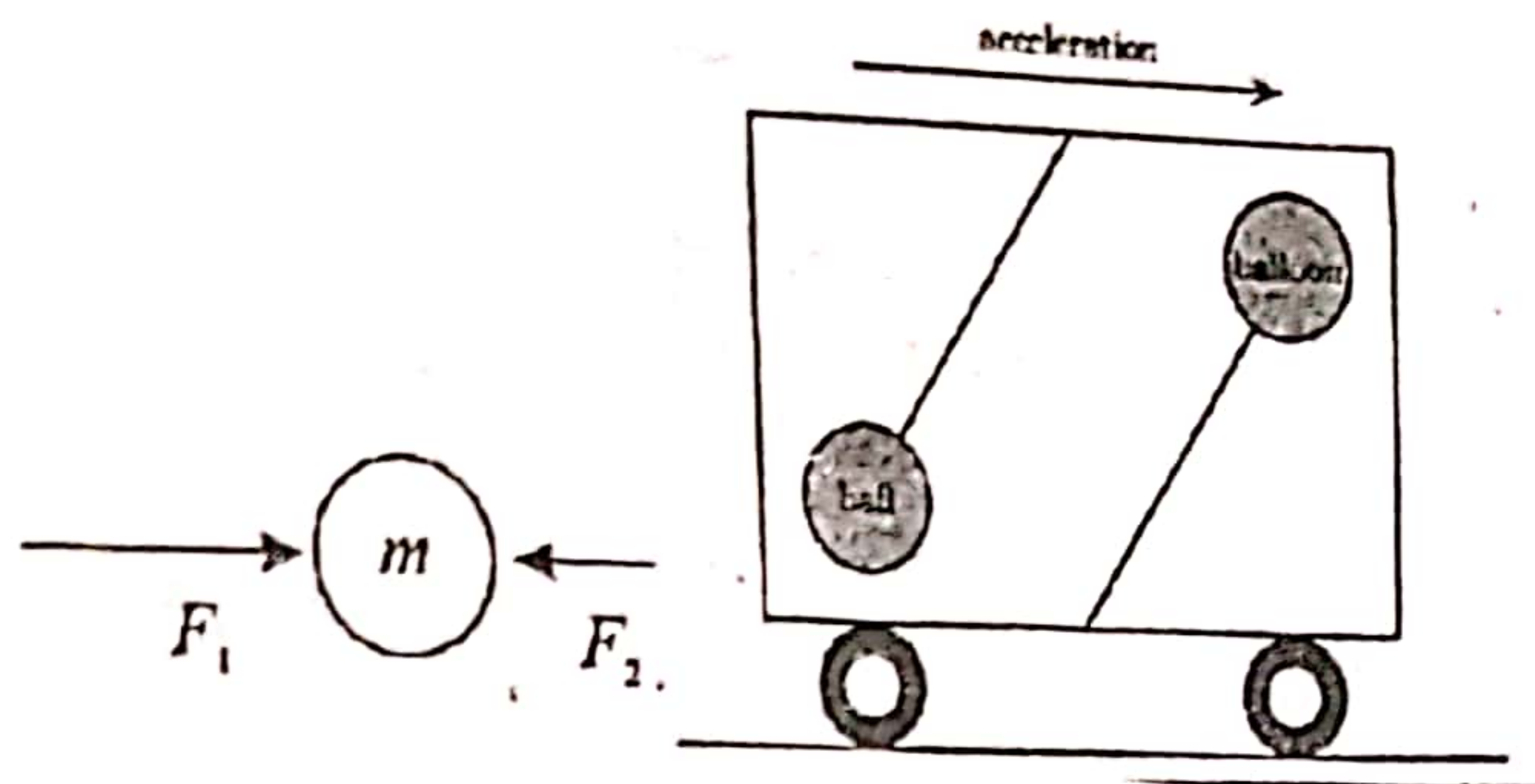
Newton's Law & Momentum

Look at the 60th question of paper 1994. There is a detailed review in it. The density of air is lesser than water. Compared to an equal volume of water mass, the mass of air volume (inertia) is less. Therefore, when the water is pressed backwards, the balloon with air goes forward. The figures show the direction of the deflection of a Helium balloon. When the container is accelerated to the left side, the air in it is pressed to the right side and the air pressure of the right side of the balloon gets increased compared to the left side air pressure of the balloon. So, the balloon filled with Helium is pushed towards the direction of the acceleration. Due to the deceleration by applying brakes, the balloon is deflected to the opposite side.

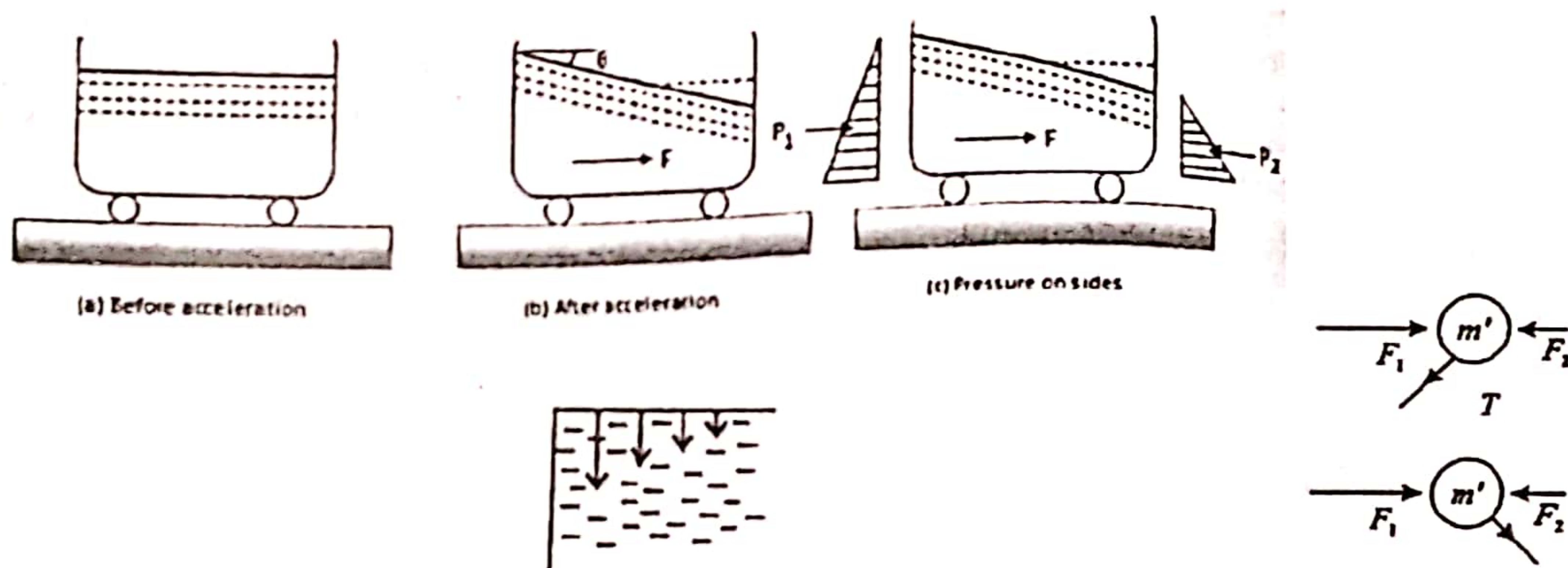


The same thing happens in water. Inside the water, there is no need to fill the balloon with Helium or a gas that is lighter than air. You can fill with normal air. That is because the normal air is lighter (less denser) than water.

When the vehicle with the water tank is accelerated to the right, the water inside the tank presses the left wall of the tank a lot. Think that there is no balloon inside the tank. Consider a water volume of mass m in the tank.



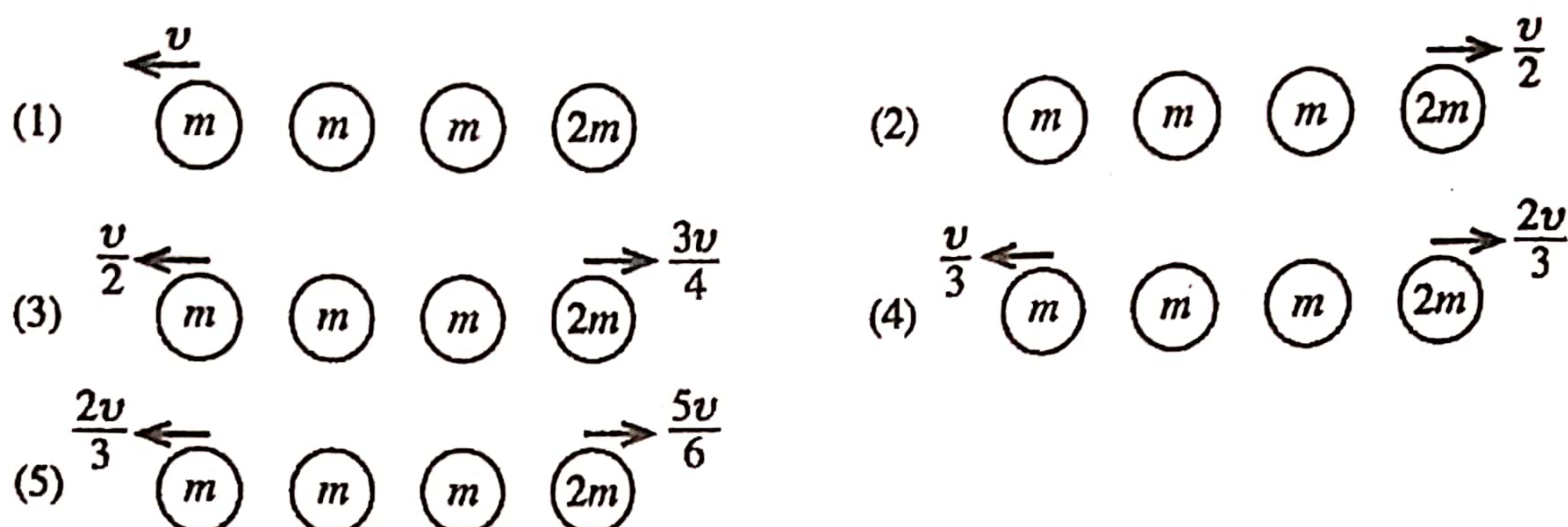
This mass is accelerated to the right side. If so, there should be resultant force on m to the right. F_1 is the outside force on m from the water pressure of left side. Likewise, F_2 is the force from the water pressure of right side. $F_1 - F_2 = ma \rightarrow F_1 > F_2$; What it means is that the water pressure from behind is greater than the water pressure from the front. If water is half filled in an open container, according to the figure, then the water on the left side will rise. If so, then you can understand that the left side pressure is greater than the right side pressure. But as the water is completely filled in the container of the question, the water of the left side will not rise. But if so, then there can be a question on how this pressure difference is created. Some have inquired about this from me.



Even the water rise from the behind is not visible, the water that is behind presses the left side of the upper lid a lot. An equal and opposite force of the vertical force created from the water on the lid is obtained from the lid to the water. So, as shown in the figure, a high pressure is applied on the water that is behind compared to water that is on the right side. At the same time, the left vertical wall of the tank is pressed more from water. So, our argument does not change as the left side water does not rise. Now instead of m water mass, let us fill it with air. If m' is the equal amount of volume for the air mass, then $m' < m$ definitely. But the values of $F_1 > F_2$ does not change as the volume of water is filled with equal volume of air. So, as $m' < m$, the air volume is pushed forward.

When the balloon is at rest relative to the vehicle, the horizontal component of the tension in the string should be towards left side for the equilibrium. As $F_1 > F_2$, $F_1 = F_2 + T \cos \theta$ can be satisfied. If it was pushed to the other side then, $F_2 = F_1 + T \cos \theta$. This can never happen. Therefore, when the vehicle is accelerated forward, the balloon should move forward. When it moves with a uniform velocity, the balloon should be directed towards the vertical direction (without a deflection) whereas when it is decelerated, the balloon should be directed backwards. Look at the 59th question of paper 2002.

36. Consider four metal balls of same volume placed on a smooth horizontal surface. Mass of each of the first three balls is m while the mass of the fourth ball is $2m$. They are equally separated on a straight line. The first ball moves with speed v and collides with the second ball resulting a series of linear elastic collisions among the balls. After all the collisions, the motion of each ball is best represented by

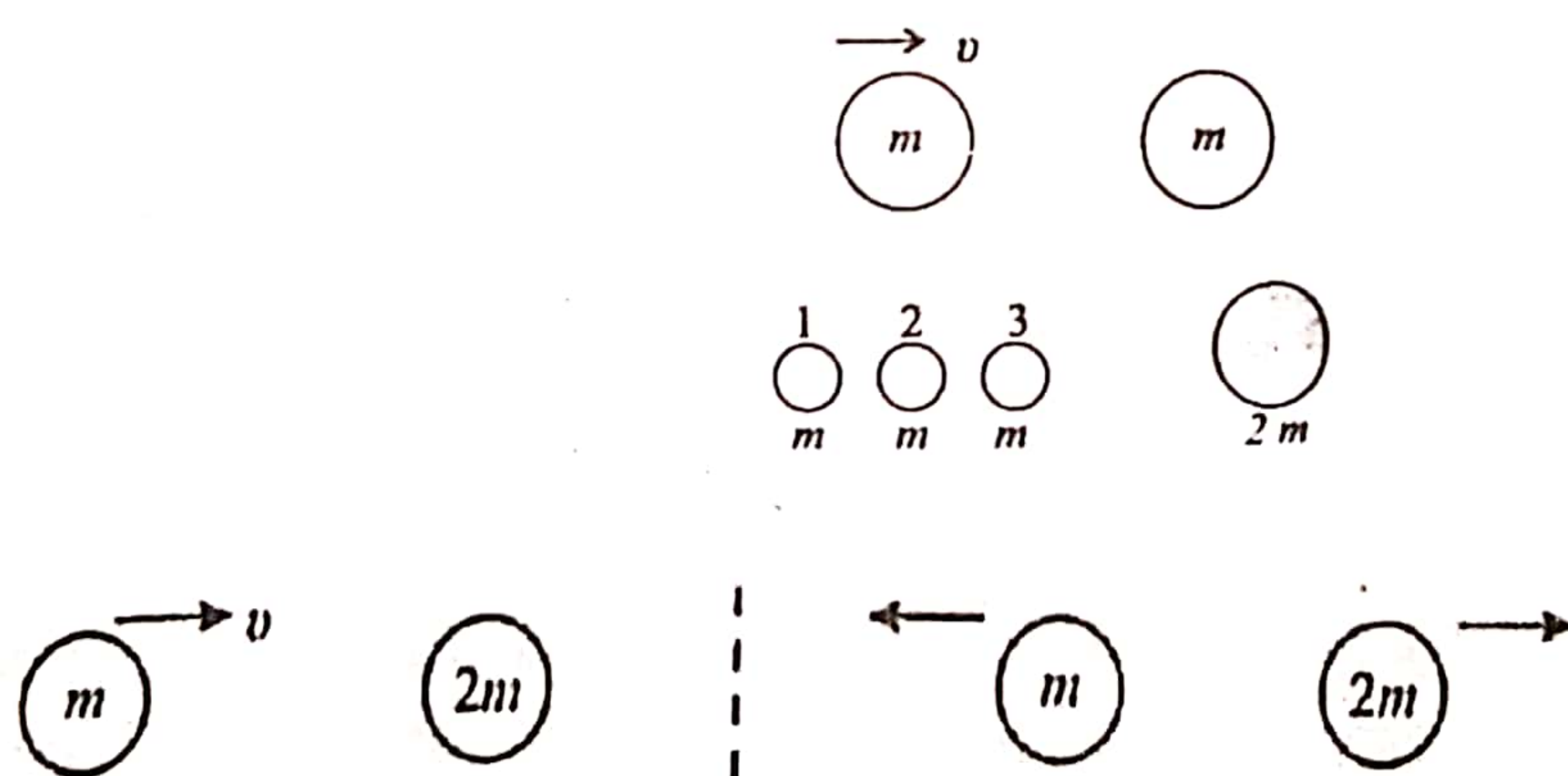


02

Newton's Law & Momentum

When two identical masses are kept at the same level and the left sphere is elastically hit with the second sphere at rest with a v speed, we know that the first sphere will be at rest while the second sphere goes forward with a v speed. This has been discussed in the 24th question of paper 2004 and the 39th question of paper 2011.

The second sphere is gone with v speed and is hit with the third sphere which is at rest. Then the second sphere will be at rest where the third sphere will go forward in v speed. This logic cannot be applied for the collision of the forth sphere. The reason is its mass is $2m$. The third sphere which comes in v speed is hit with $2m$ and bounced back. That is because $2m$ is bigger than m . When you collide with a person with a greater mass, normally bouncing back occurs. $2m$ will go forward at a certain speed. The third sphere will turn back in a certain speed. It will collide with the second sphere, gives its speed to the second and comes to rest. The second sphere will collide again with the first one and comes to rest. The first sphere will move to left side with the speed of the second sphere. However, in all four answers the middle two masses are given at rest. From that we do not get a hint to reach the correct answer.



Actually, even there are four spheres, you need to consider only the collision of the third sphere (m) with the fourth sphere which is at rest.

The fourth sphere goes forward at a certain speed. The third sphere turns back and gives its speed finally to the first one.

First, apply conservation of momentum. Initially, the momentum of the system is $\rightarrow mv$. Therefore, whatever happens, the final momentum also should be $\rightarrow mv$. From that, the first choice is removed. Its final

momentum is $\leftarrow mv$ So, remove that. The other point is that $2m$ cannot be at rest.

Even the net momentum of the second choice $\rightarrow mv/2 = \rightarrow mv$ the third sphere bounces back and gives its speed to first m . So, the first sphere cannot be at rest. So, remove that also.

In the third choice, let us consider the net momentum. I will not write m and v hereafter. The resultant momentum = $\rightarrow 2 \times \frac{3}{4} - \frac{1}{2} = 1 \left[2m \cdot \frac{3}{4}v - \frac{mv}{2} = mv \right]$. This is correct from conservation of momentum. But do not select this as the correct answer. As it is an elastic collision, the kinetic energy should also be conserved.

The initial kinetic energy is $\frac{1}{2}mv^2$. Therefore, the final kinetic energy (of the system) should be $\frac{1}{2}mv^2$. The final kinetic energy of the third choice = $\frac{1}{2} \cdot \frac{1}{4} + \frac{1}{2} \cdot 2 \cdot \frac{9}{16} = \frac{1}{8} + \frac{9}{16}$. This is not equal to $\frac{1}{2}$. $\left[\frac{1}{2}mv^2/4 + \frac{1}{2} \cdot 2m \cdot 9v^2/16 = m/2 (1/4 + 9/8)v^2 = \frac{1}{2}m \cdot 11/8 v^2 \right]$

Now let us go for the fourth choice. Net momentum = $2 \times 2/3 - 1/3 = 1$. It is correct. Kinetic energy = $\frac{1}{2} \cdot 1/9 + \frac{1}{2} \cdot 2 \cdot 4/9 = \frac{1}{2} (1/9 + 8/9) = \frac{1}{2}$. It is correct. The correct answer is (4). Do not look at (5). You can check it later (once you go home). It takes times as we need to check both conservation of momentum and kinetic energy. But there is another shorter method.

Shortest Method

Let us consider the masses of first and second objects as m_1 and m_2 . Even we will consider the initial velocities of the first and second objects as u_1 and u_2 whereas the final velocities will be v_1 and v_2 respectively. If the collision is elastic, then from the conservation of kinetic energy,

$$\frac{1}{2} m_1 u_1^2 + \frac{1}{2} m_2 u_2^2 = \frac{1}{2} m_1 v_1^2 + \frac{1}{2} m_2 v_2^2 \dots (1)$$

From the conservation of momentum, $m_1 u_1 + m_2 u_2 = m_1 v_1 + m_2 v_2 \dots (2)$

From the equation (1), $m_1 (u_1^2 - v_1^2) = m_2 (v_2^2 - u_2^2) \rightarrow m_1 (u_1 + v_1)(u_1 - v_1) = m_2 (v_2 + u_2)(v_2 - u_2) \dots (3)$

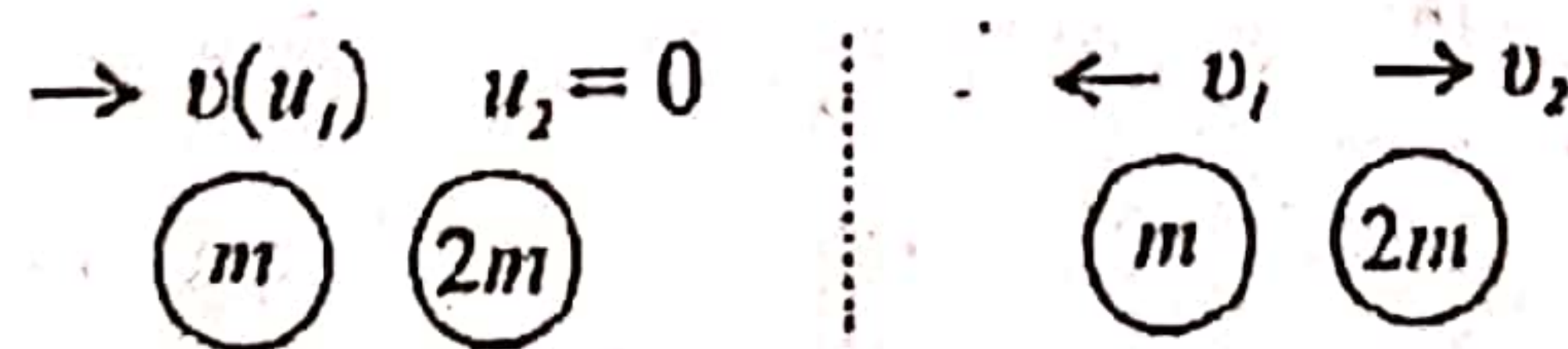
From the equation (2), $m_1 (u_1 - v_1) = m_2 (v_2 + u_2) \dots (4)$

$$(3)/(4) \quad u_1 + v_1 = u_2 + v_2$$

If you remember this, then you can get the answer quickly. It is easy to remember as well.

The initial velocity of the first object + the final velocity of the first object = the initial velocity of the second object + the final velocity of the second object.

This equation does not have masses. You need to keep in mind that this can be applied to an elastic collision. Also remember that the velocity is a vector quantity. Now we will apply for our question. As mentioned earlier, even there were three collisions, you need to consider only the final collision.



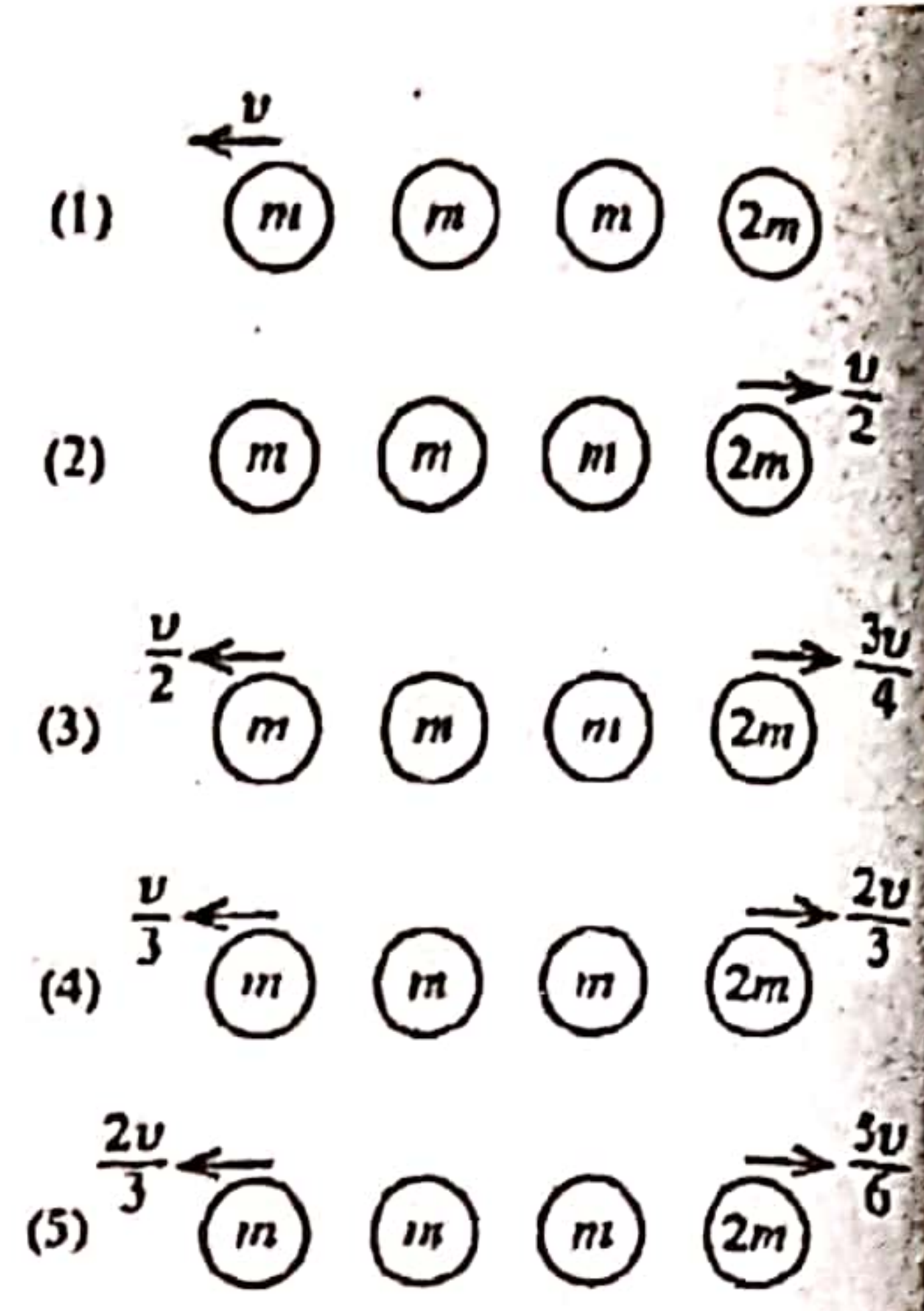
The choices of (1) and (2) can be just removed. Check by applying the answers of other choices.

Choice (3) $v - v/2 = 0 + 3/4 v \rightarrow$ It is not matching.

Choice (4) $v - v/3 = 0 + 2v/3 \rightarrow$ It matches perfectly.

Choice (5) $v - 2v/3 = 0 + 5v/6 \rightarrow$ It does not fit.

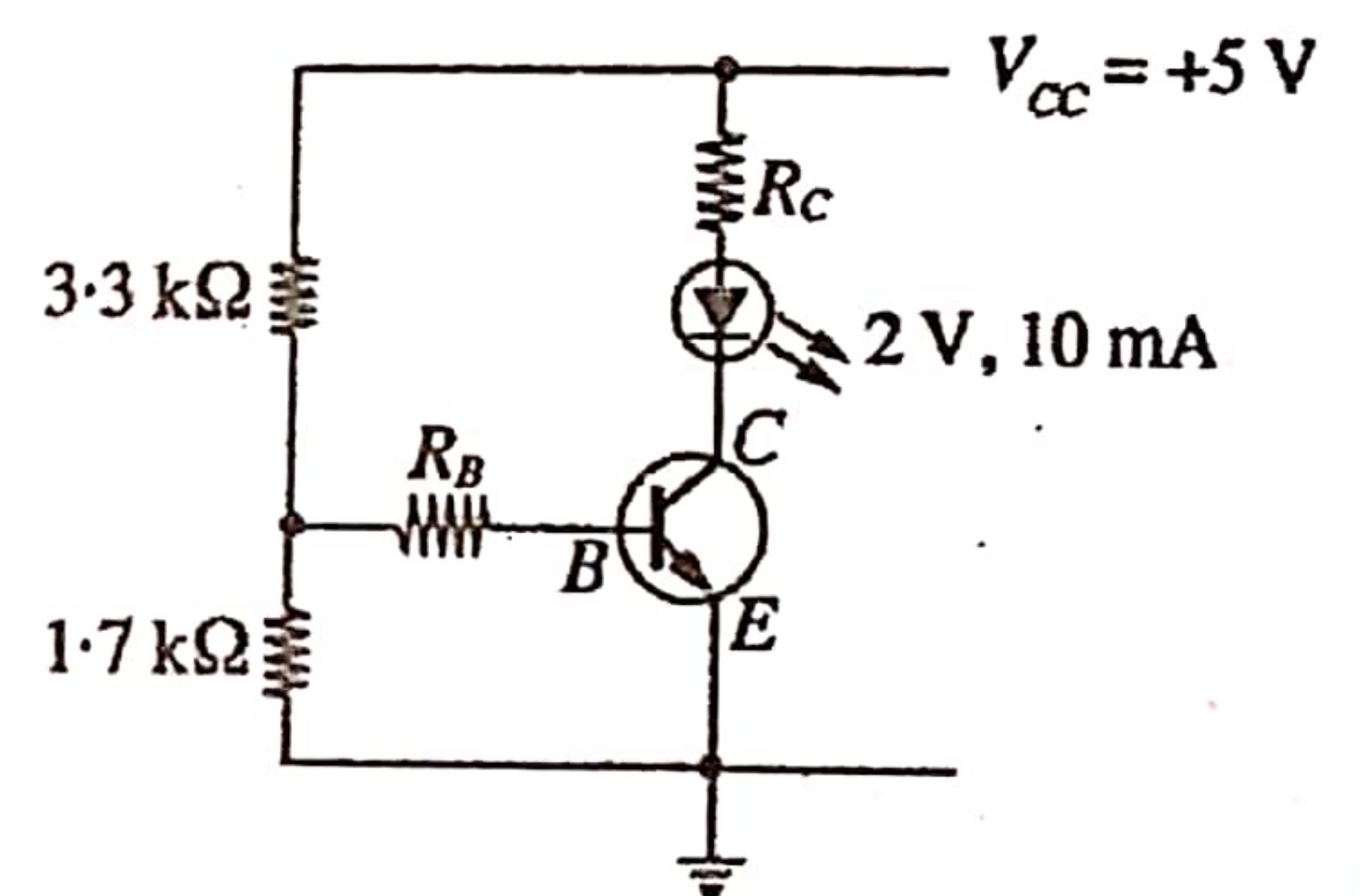
In a normal problem, we do not know both of v_1 and v_2 . Therefore, from the above equation only we can not find both v_1 and v_2 . So, from (1) and (2) equations we can get expressions for v_1 and v_2 . In this MCQ, as v_1 and v_2 are given, we can substitute and check them quickly.



The above expression can be expressed as $v_1 - v_2 = -(u_1 - u_2)$. This fairly represents in applied mathematics as $v_1 - v_2 = -e(u_1 - u_2)$. Here, $(v_1 - v_2)$ = the relative velocity after the collision and $(u_1 - u_2)$ = the relative velocity before the collision. This is not in the syllabus of Physics. This is called as 'Newton's Law of Restitution'. It is a simple law where e is known as the coefficient of restitution. For an elastic collision $e = 1$.

37 For optimum operation of a light emitting diode (LED), forward voltage and current should be 2 V and 10 mA, respectively. Transistor is having $V_{BE} = 0.7$ V, current gain $\beta = 100$, and $V_{CE(sat)} = 0.1$ V. For the circuit shown in the figure, what are the values of R_B and R_C for the optimum operation of the LED?

- (1) $R_B = 100 \Omega$ and $R_C = 1 \text{ k}\Omega$
- (2) $R_B = 1 \text{ k}\Omega$ and $R_C = 1 \text{ k}\Omega$
- (3) $R_B = 1 \text{ k}\Omega$ and $R_C = 290 \Omega$
- (4) $R_B = 10 \text{ k}\Omega$ and $R_C = 1 \text{ k}\Omega$
- (5) $R_B = 10 \text{ k}\Omega$ and $R_C = 290 \Omega$



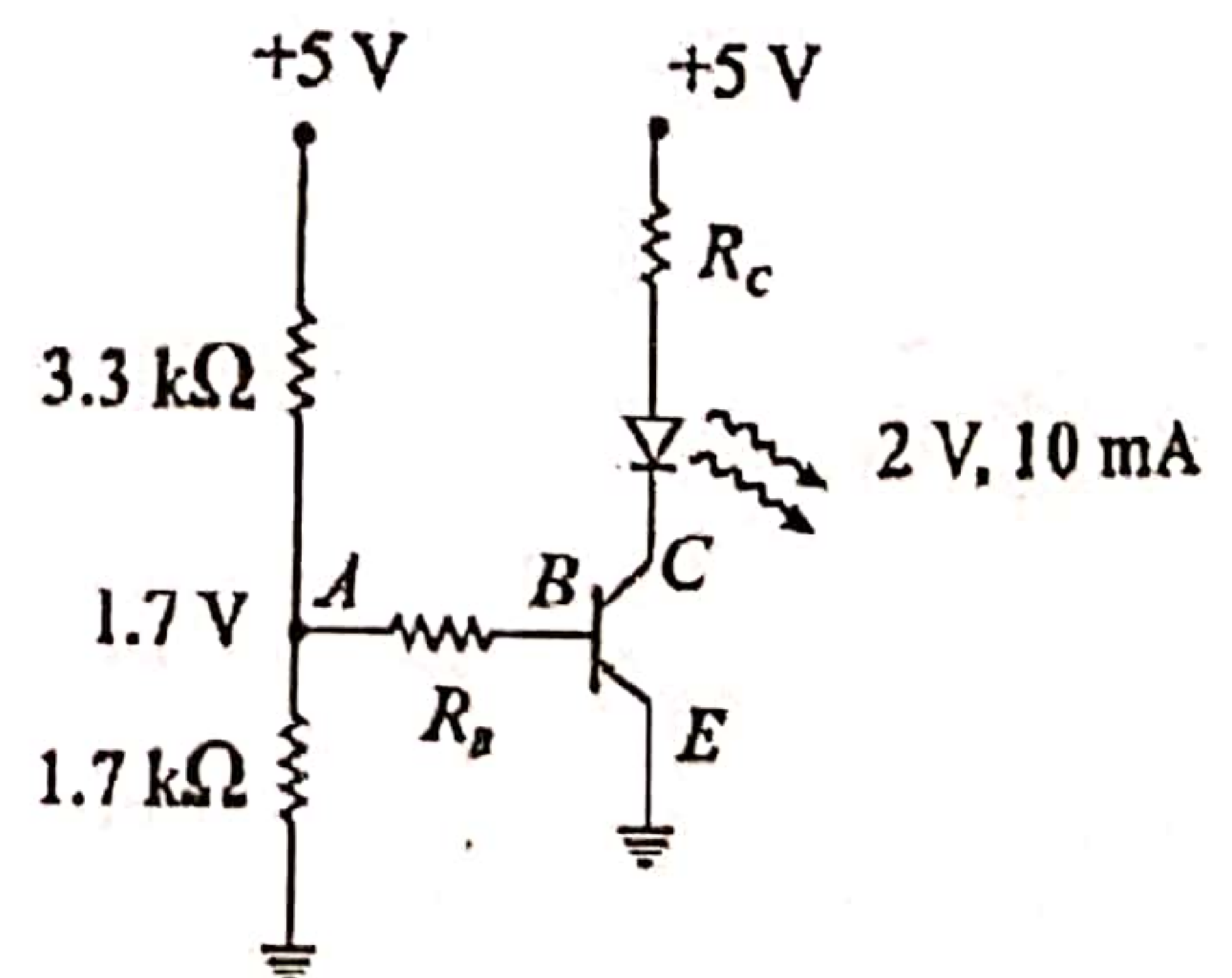
09

Transistors

The values are given for the optimal performance of the LED. Actually, it is easy to find R_C first. There should be a 2V voltage drop across the LED. As V_{CE} is given as 0.1 V, the potential drop across $R_C = 5 - (2 + 0.1) = 2.9$ V. The current that should flow across LED flows across R_C too. So,

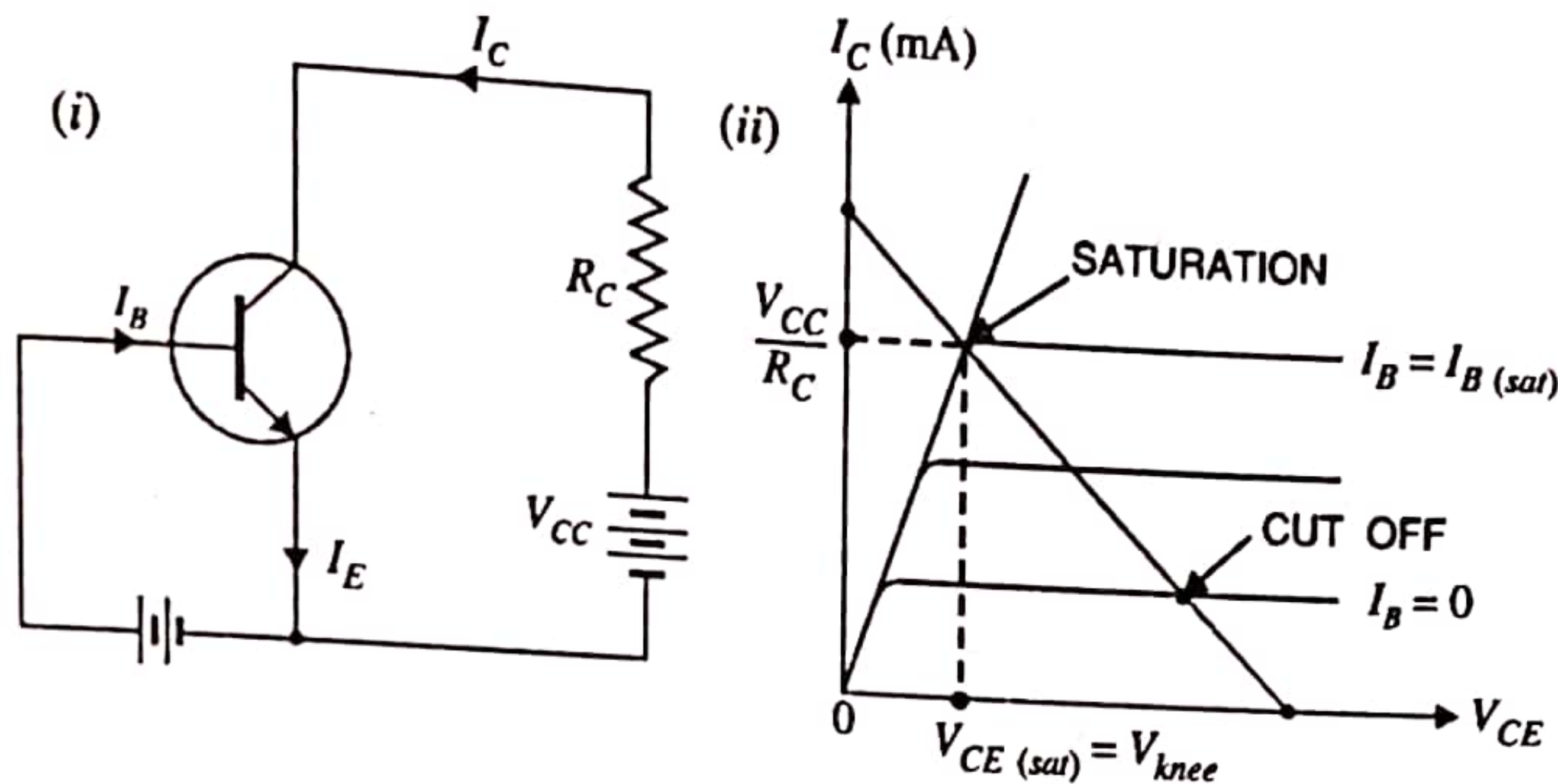
$R_C \times 10 \times 10^{-3} = 2.9 \rightarrow R_C = 290 \Omega$. Next, to find R_B , we need to find V_{AB} (the voltage drop across R_B) and I_B . +5 V is divided among 3.3 kΩ and 1.7 kΩ. As the total of (3.3 + 1.7) is exactly 5, 5V is nicely divided as 3.3 V across 3.3 kΩ and 1.7 V across 1.7 kΩ. As the lower end of 1.7 kΩ resistor is earthed, $V_A = 1.7$ V. The emitter of the transistor is also directly earthed.

So, $V_{AB} = 1.7 - 0.7 = 1$ V. To find I_B , use $\beta = I_C/I_B$. $I_B = I_C/\beta = (10$



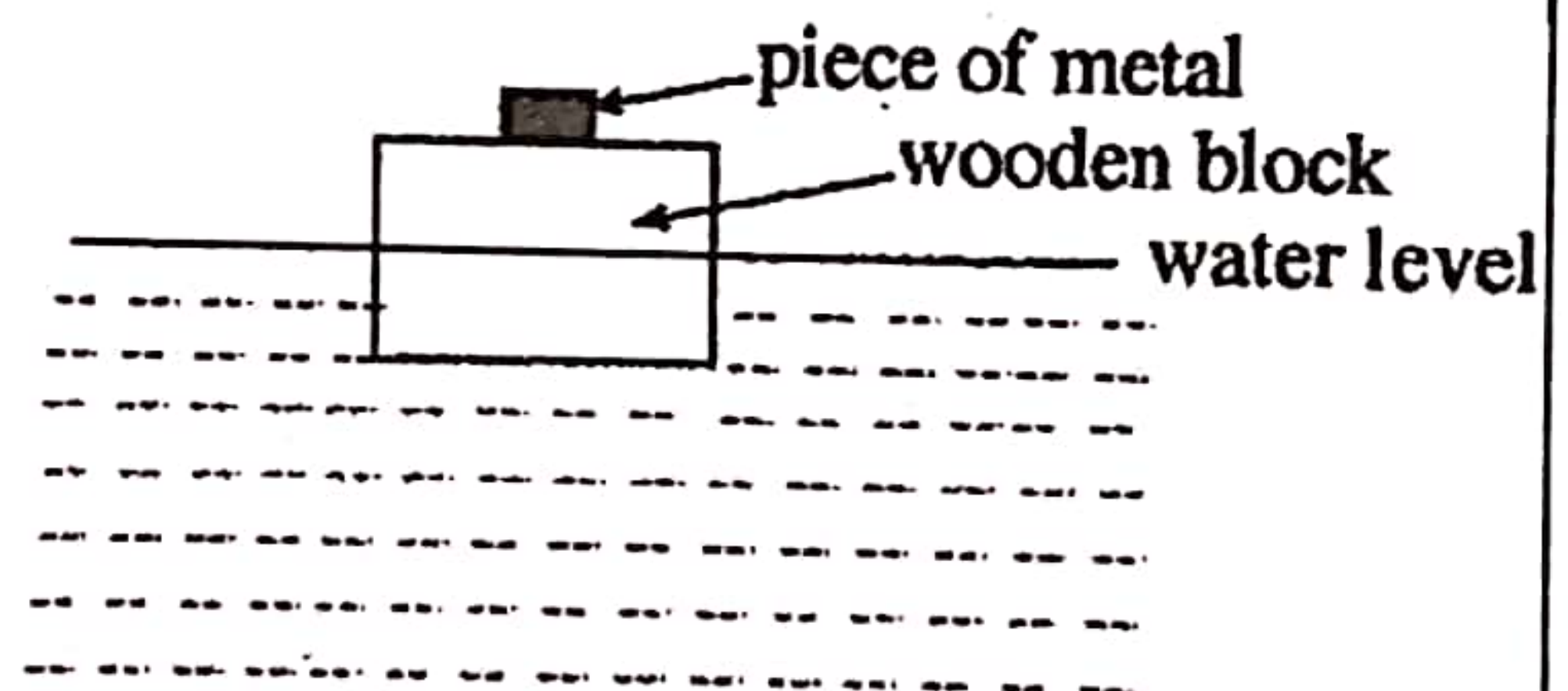
$\times 10^{-3})/100 = 10^{-4}$ A. Now apply $V = IR$ across R_B , $1 = R_B \times 10^{-4}$, $R_B = 10^4 \Omega = 10 \text{ k}\Omega$; $R_B = 10 \text{ k}\Omega$ and $R_C = 290 \Omega$.

As $V_{CE \text{ saturated}} = 0.1 \text{ V}$ is given, there can be a problem about whether we can apply $\beta = I_C/I_B$ for the transistor. That is because there can be a doubt about whether the transistor is existing in the active region. But you do not have to make this as an issue.



You can consider that the transistor is exactly at the corner of the active region (look at the figure). You can apply $\beta = I_C/I_B$ for that place. This relation cannot be applied afterwards. That means when it comes to saturated region. However, this question cannot be solved without giving V_{CE} or without finding I_B value. Therefore, we need to argue based on that foundation. So, there is no other alternative for us than using the given data.

38. A piece of metal is attached to the top of a rectangular wooden block that floats in water. As shown in the figure, 50% of the volume of the wooden block is submerged in water. The metal piece and the wooden block have the same mass. If the wooden block with the metal piece is flipped up side down, what could be the percentage of the volume of the wooden block submerged in water?



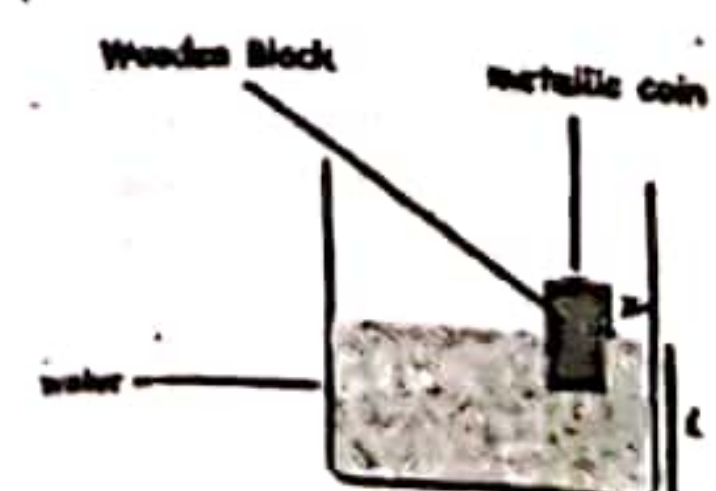
- (1) Slightly smaller than 50% (2) Much smaller than 50% (3) 50%
(4) Slightly larger than 50% (5) Much larger than 50%

Hydrostatics

02

A metal coin is fixed on a wooden block. According to the figure, 50% of the volume of the wooden block is kept in the water and it floats on the water. If the wooden block is turned upside down and sunk in water, then the volume that the wooden block is sunk will be lesser than before. This is general knowledge. Initially, the wooden block is sunk in the water. The weight of the system does not change by changing the block into upside down position. Now as the coin also sinks in water, there will be an upthrust on it too.

Therefore, definitely the volume of the wooden block that sinks in water should be less. Initially, the coin is not sunk. Therefore, the upthrust equal to the weight is given only by the wooden block. Next, as the coin also sinks in water, there is an upthrust from it too. The weight of the compound has not been changed. Therefore, the volume of the wooden block that sinks in water should be definitely be less.



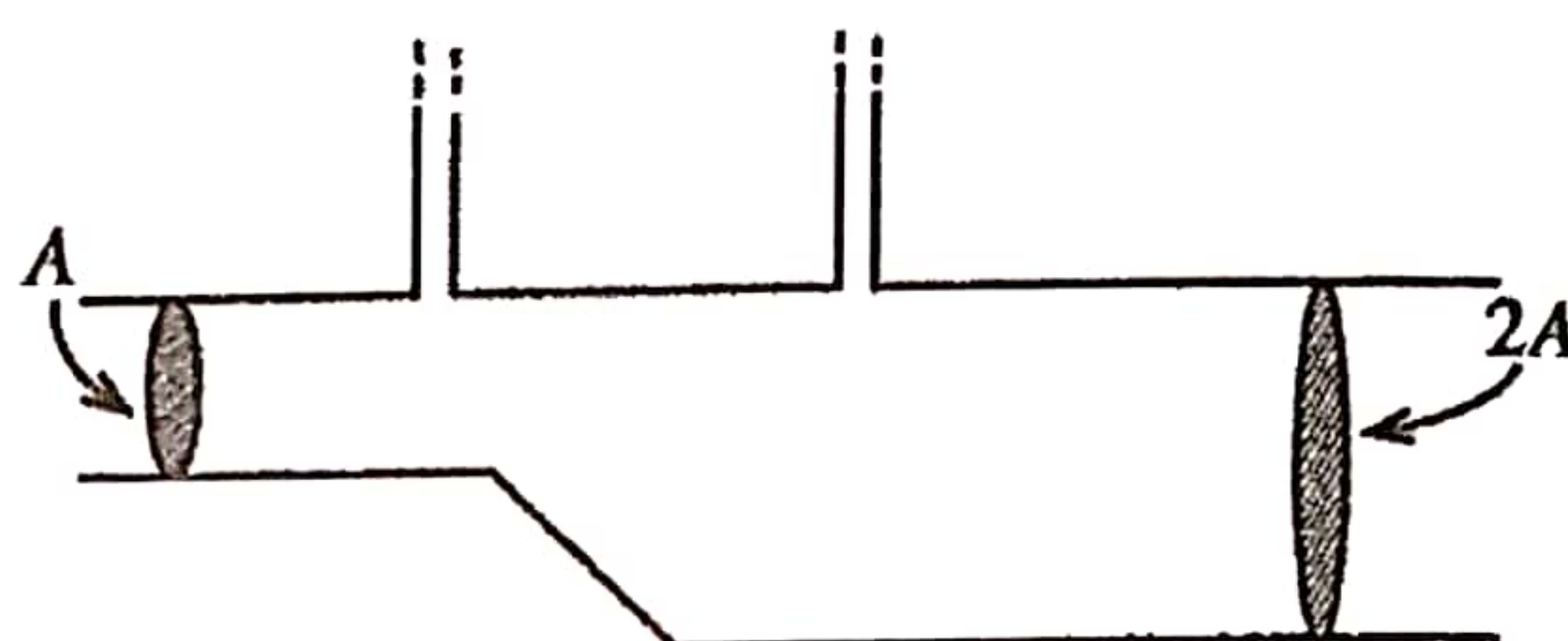
How much will it be reduced? If the wooden block and the coin has equal masses, then as the metal's density is higher than the wood's density, the volume of the metal is lesser than the volume of the wooden block. Even from the figure it can be seen. Therefore, the upthrust from the metal is not in a bigger range. So, the

volume that the wooden block sinks in water is lesser than 50%.

First, you need to decide that the volume of the wooden block that sinks in water should be lesser. Then only two choices are left. To decide whether the wooden block is sinking less or slightly, you need to compare the volumes of the wooden block and the metal. Similar masses of wooden block and the metal have been given to do the volume comparison.

39. As shown in the figure, an incompressible liquid flows steadily through a horizontal pipe. Two narrow vertical tubes are fixed at two places on the horizontal pipe where the cross-sectional areas are A and $2A$. If the height difference of the liquid columns in the two vertical tubes is h , flow rate of the liquid through the pipe is

- (1) $A\sqrt{2gh}$ (2) $A\sqrt{6gh}$
 (3) $A\sqrt{\frac{3gh}{2}}$ (4) $2A\sqrt{\frac{gh}{3}}$
 (5) $2A\sqrt{\frac{2gh}{3}}$

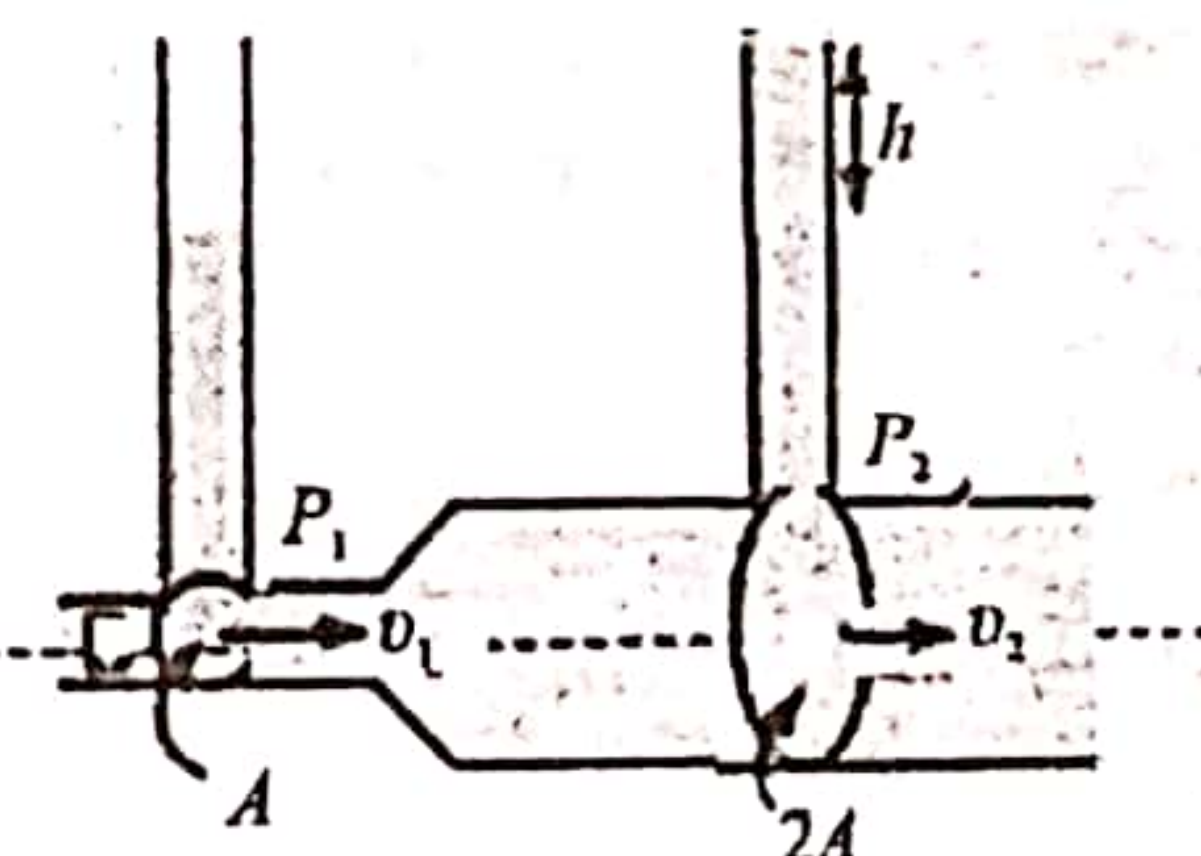


02

Hydrodynamics

This is a question which applies Bernoulli's principle to solve. Consider the venturimeter as shown in the figure.

This is being used to measure the speed of fluid flow. When the tube which the fluid flows gets narrower, then the speed of the fluid gets increased. Then its pressure is reduced. When the tube is wider, then the speed gets reduced and the pressure gets increased. This pressure difference can be measured from the height differences of the fluids that are connected as vertical tubes. If Bernoulli's equation is applied to the flow line shown in the dashed line, then $h\rho g = \frac{1}{2}\rho v_1^2 - \frac{1}{2}\rho v_2^2$. In addition, $Av_1 = 2Av_2 \rightarrow v_2 = \frac{1}{2}v_1$



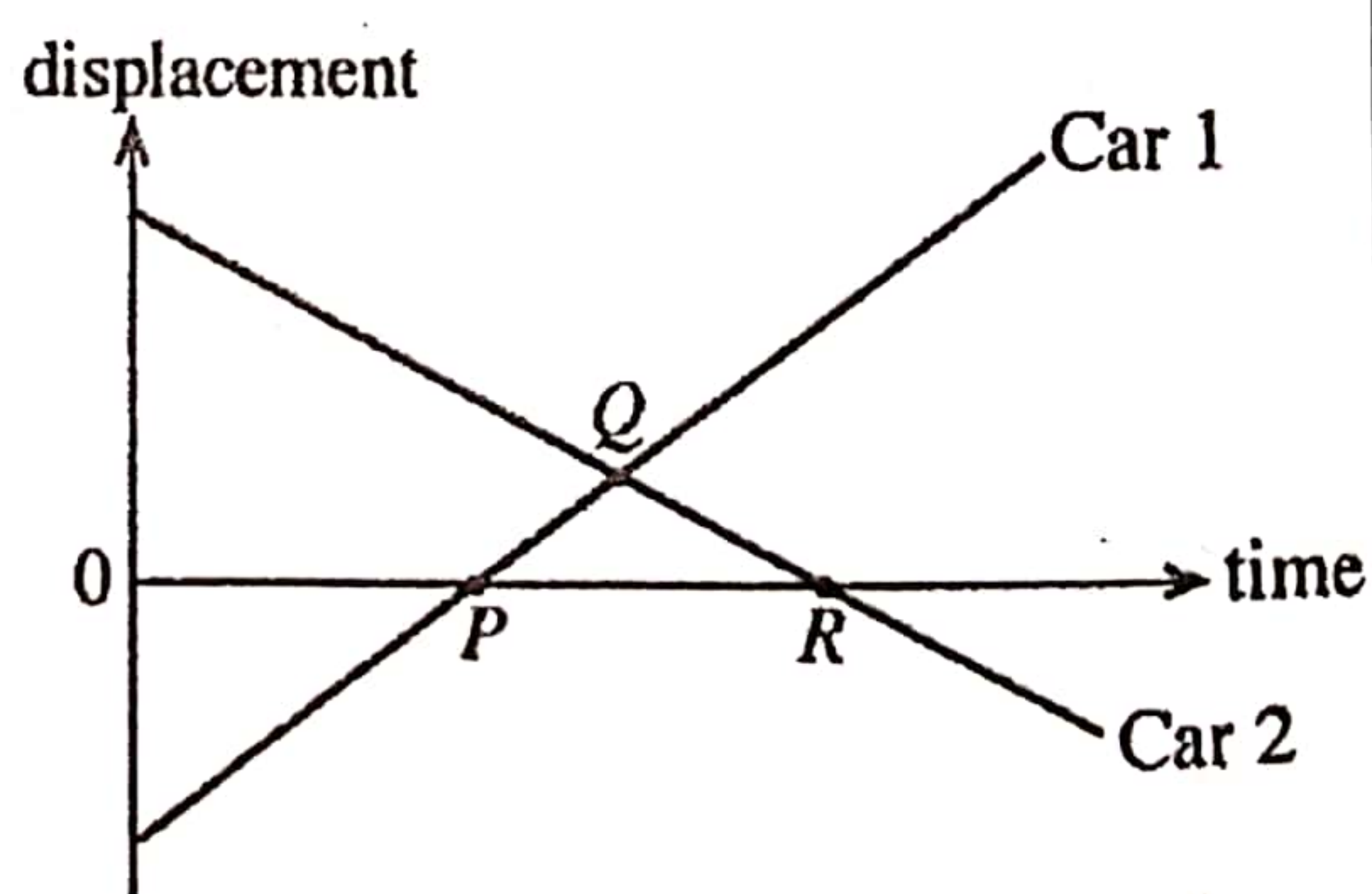
Substituting v_2 into first equation, $2gh = v_1^2 - \frac{1}{4}v_1^2 = \frac{3}{4}v_1^2 \rightarrow v_1^2 = \frac{8gh}{3}$; $v_1 = 2\sqrt{\frac{2gh}{3}}$

40. The figure shows the displacement-time graphs for the motion of two cars with respect to a lamp post aside the road. Consider the displacement to the right side of the lamp post as positive. A student has made the following statements regarding the motion of cars relevant to the points P , Q , and R marked on the graph.

- (A) Relevant to P : Car 1 coming from left crosses Car 2.
 (B) Relevant to Q : Both cars are moving towards the lamp post and cross each other.
 (C) Relevant to R : Car 2 coming from right passes the lamp post.

Which of the above statements is/are correct?

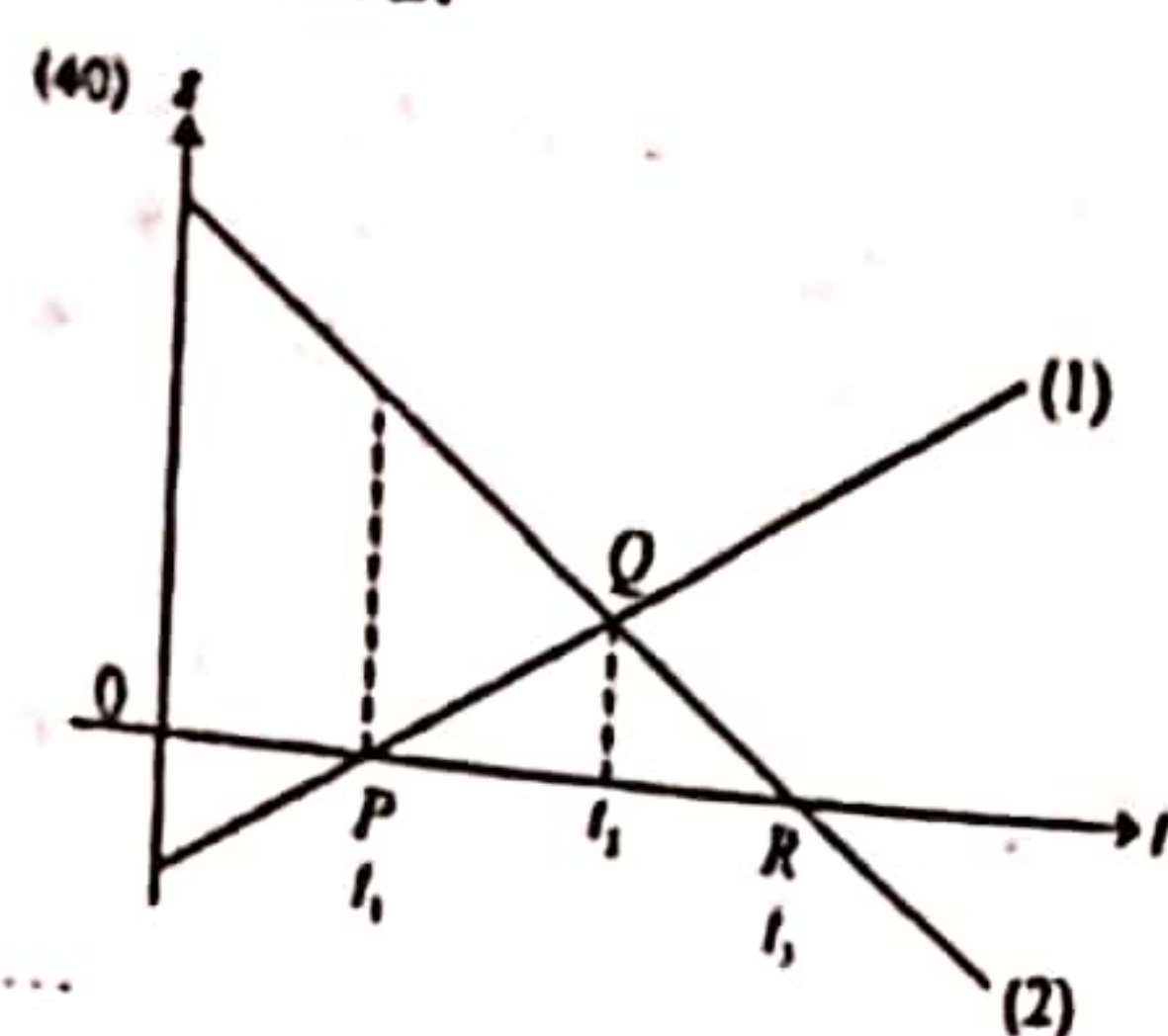
- (1) Only B (2) Only C (3) Only A and B
 (4) Only B and C (5) All A, B, and C



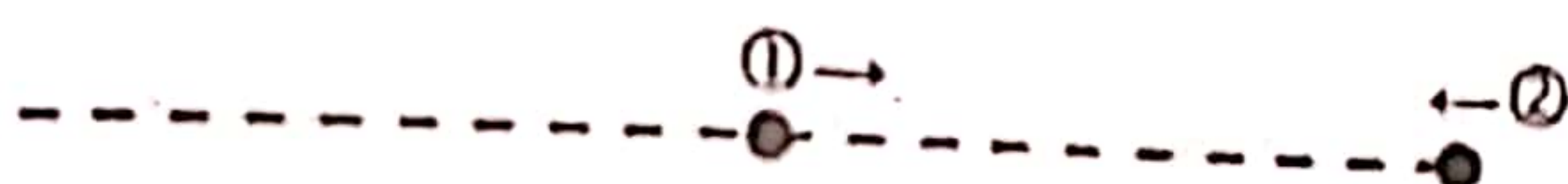
02


Linear Motion

The displacement-time curves of two vehicles are shown in the figure. According to the graphs, the points where the displacement get zero are P and R. The first vehicle starts from the left side of the point where the displacement is zero (point where the displacement is measured). Then it passes that point (when $t = t_1$) and then it goes to the right side. The second vehicle starts from the right side and go towards the point which has a zero displacement. Then it passes that point (when $t = t_1$) and go to the left side. You can get the correct answer easily if you put these facts into your head.

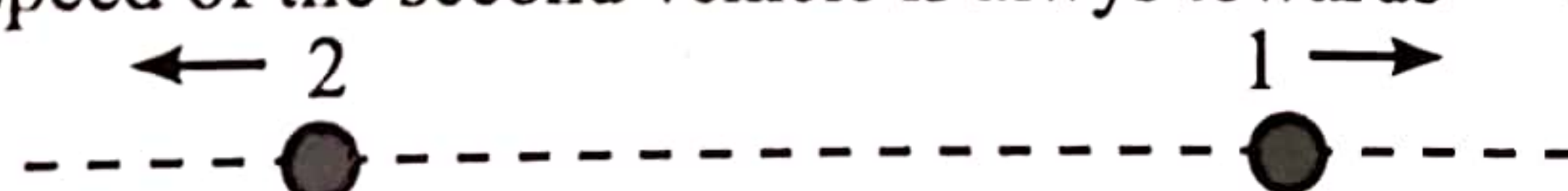


- (A) When $t = t_1$ (at P), the vehicle that is coming from the left side passes the origin (the zero displacement point). But the second vehicle is to the right side of the origin (as there is a positive displacement). Therefore, there is no chance that these two vehicles can pass each other. If needed, then we can show the places of the vehicles when $t = t_1$ like this way. This sentence is false.



- (B)  When $t = t_2$ (at Q), the displacement of both vehicles are positive and equal to each other. That means they pass with each other. But they are going to two sides. The first vehicle is going away from the origin whereas the second vehicle is going towards the origin here. This situation can be represented like this way. Even they are passed with each other, they go towards different directions. Hence, this sentence is also false.

- (C) When $t = t_3$, the second vehicle passes the origin and moves to the left side. This is true. This situation can be represented like this way. This sentence is only correct. The gradient of the s-t curve of the first vehicle is positive. That means the speed of the first vehicle is always towards \rightarrow direction. Likewise, the speed of the second vehicle is always towards \leftarrow direction.



41. A whistling firecracker having a constant whistling frequency is fired vertically upward. It travels initially with an acceleration, then with a deceleration, and finally blasts before coming to the rest. An observer at ground directly below the firecracker listens to the whistling sound of the firecracker.

Consider the following statements regarding the frequency of the sound heard by the observer.

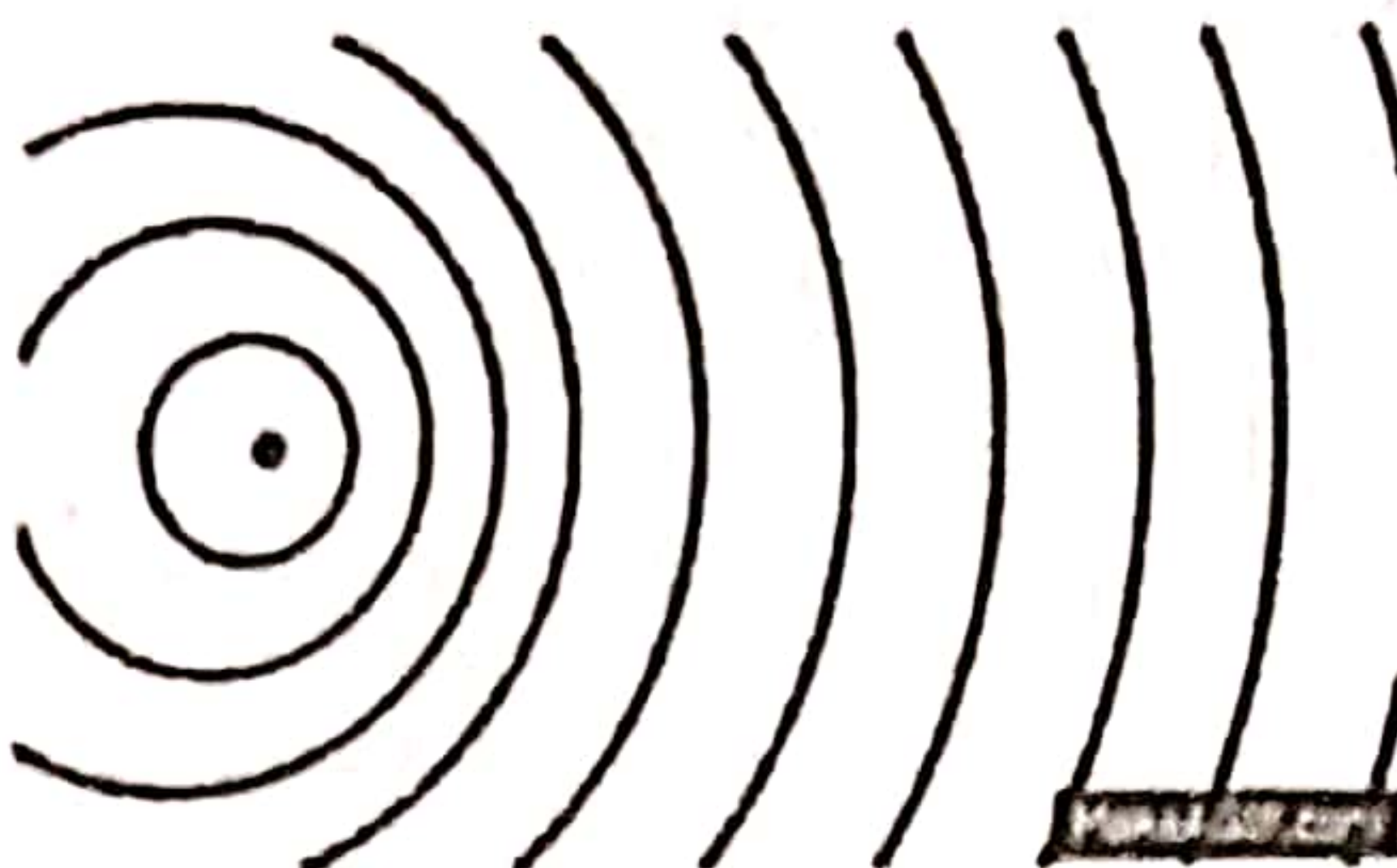
- (A) During the acceleration, it is higher than the whistling frequency and is decreasing with time.
 (B) During the deceleration, it is lower than the whistling frequency and is increasing with time.
 (C) Just before the blast, it becomes equal to the whistling frequency. Which of the above statements is/are correct?

- (1) Only A (2) Only B (3) Only C
 (4) Only A and B (5) Only B and C

Doppler Effect

03

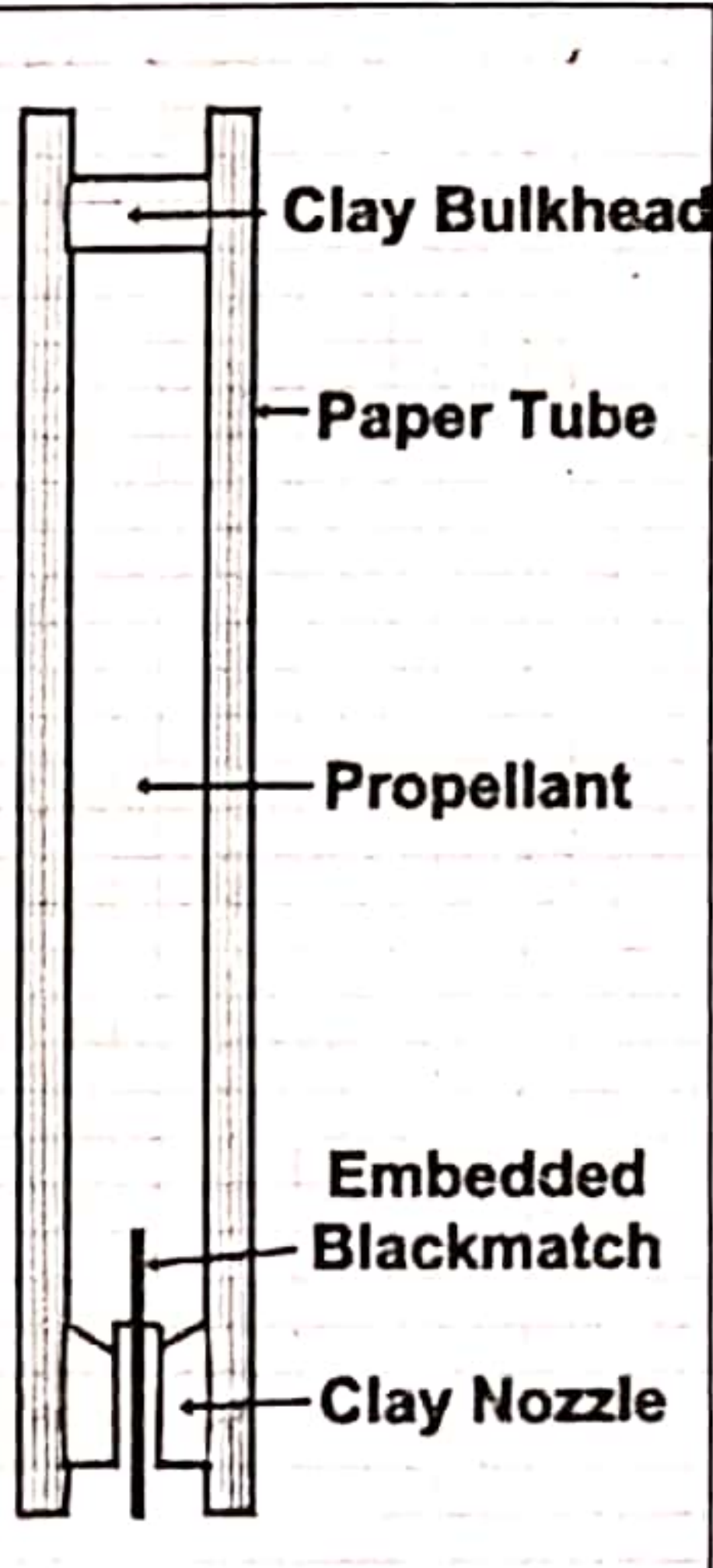
This is a beautiful question. Whether the firecracker is accelerated or decelerated, it goes away from the observer; That means the direction of its speed is away from the observer. It is not towards to the observer. If so, then the apparent frequency should be clearly lesser than the true frequency of the horn. It cannot be greater. From that, the first sentence becomes false.



Half of the second sentence is true at a glance. The apparent frequency is small. As it is getting decelerated, the speed away from the observer is gradually getting reduced. It slows down the getting away. Therefore, even the apparent frequency is small, it increases gradually with the time. Think that it decelerated and came to rest. Then apparent frequency of that instance should be equal to the true frequency. If you think in a way that the reduced apparent frequency should be equal with the true frequency, then it should be increased with the time.

As it is given that the explosion occurs before it is at rest, then the frequency heard before the explosion cannot be equal to the horn frequency what so ever. The apparent frequency at a moment before the explosion or at any time before that is smaller than the true frequency. The frequencies get equal only if the firecracker is at rest.

As long as the gases are emitted from the bottom by the burning of the chemicals (gun powder) in the firecracker, it is accelerated (like a rocket). The gases emitted from the burn give an upward pull to the firecracker. The firecracker gets decelerated when the chemicals are completely burnt out. When the chemicals are burnt out, then the pull diminishes. If we neglect the upthrust and the air resistance, then the firecracker is subjected to a deceleration of g .



Why do we hear a sound of a whistle like from a horn when some chemicals are burnt in a firecracker? Look at the figure. When the burnt gas is emitted from a nozzle with a small hole in a tube, a similar sound corresponding to the whistle sound of a horn can be heard.

42. A metal bowl of mass 700 g contains 1 litre of water at 27 °C. When a steel ball of mass 300 g at 120 °C is dropped into the water in the bowl, the final temperature of water is measured to be 30 °C. Specific heat capacities of steel and water are 500 I kg⁻¹ K⁻¹ and 4200 J kg⁻¹ K⁻¹ , respectively. Out of the metals given in the table, what could be the metal that the bowl is made of?

- (1) Aluminium

(2) Copper

(3) Lead

(4) Iron

(S) Silver

Metal	Specific Heat Capacity (J kg ⁻¹ K ⁻¹)
Aluminium	900
Iron	450
Copper	385
Silver	230
Lead	128

A calculation is needed and it takes some time. The steel ball gives the heat. The container and water absorb the heat. 1 l of water means 10^3 cm^3 . The density of water is 10^3 kgm^{-3} . That means 1 gcm^{-3} . That means 1 l of water is 1 kg (10^3 cm^3 is $10^3 \text{ g} = 1 \text{ kg}$). If you do not know these things, then you cannot solve this problem. I feel that instead of the volume of water, the mass should have been given. These things take time. If the specific heat capacity of the container is s , then

$$300 \times 10^{-3} \times 500 (120-30) = 700 \times 10^{-3} s (30-27) + 1 \times 4200 (30-27)$$

$$150 \times 90 = 2.1 s = 12600 \rightarrow 2.1 s = 13500 - 12600 = 900 \rightarrow s = 900/2.1 = 429 (428.6) \text{ J kg}^{-1}\text{K}^{-1}$$

Such questions with simplifications cannot be given. There is no such a metal matching to 429. If we take 2.0 instead of 2.1, $s = 450 \text{ J kg}^{-1}\text{K}^{-1}$. Many children will pick iron no doubt. Here why has copper been selected for the metal? Is it due to the fact that there are copper calorimeters?

The above calculation was done without considering the heat loss. If we consider the heat loss, then a certain amount of heat should be added to the right side of the equation. The heat loss should be added not to the person who gave the heat but to the person who got the heat. If so, then s should get a value less than 429. According to the given values, 385 is the nearest value of 429 and a lesser value. Therefore, copper has been selected. As 230 and 128 are also lesser than 429, one can argue that why cannot we select those as well. At a glance, there is a certain validity in that. But when the values of the specific heat capacities get smaller and smaller, the amount of heat loss is greater.

For an example, if $s = 385$ is taken, then the amount of heat that the container absorbed $= 2.1 \times 385 = 808.5 \text{ J}$. The amount of heat loss to the environment $= 13500 - (808.5 + 12600) = 91.5 \text{ J}$.

This value is lesser than the amount of heat that the container absorbed (808.5 J) $91.5 < 809$. But if $s = 230$, then the amount of heat that the container absorbed $= 2.1 \times 230 = 483 \text{ J}$. The heat loss to the environment for this moment $= 13500 - (483 + 12600) = 417 \text{ J}$; This amount of heat same as the amount that the container absorbed. 483 and 417.

If lead is selected for the container, then the heat loss to the environment is greater than the absorbed heat by the lead container. Normally, we expect to keep heat loss to the environment at a minimum level. In that logic, there is no wrong in taking the nearest value lesser than 429. But all these arguments we can show by staying at home. There is no time for children to think such arguments like this way. Therefore, here are the arguments that I present.

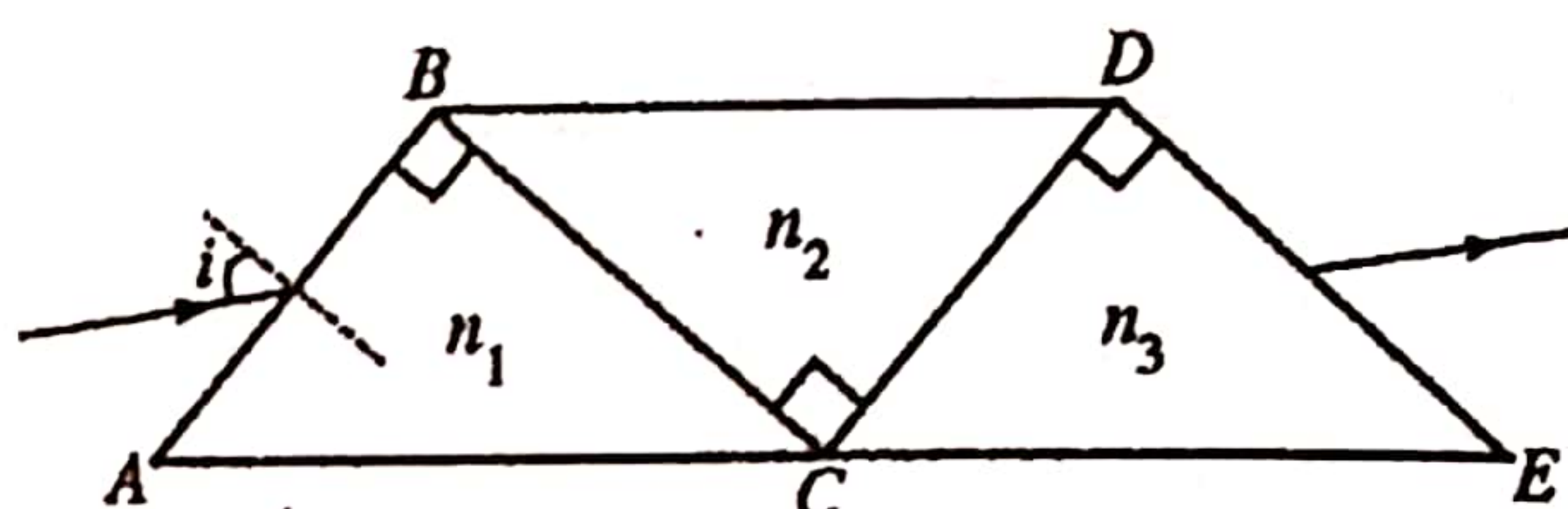
- (1) However, you need to do the question. Then you will get $429 \text{ J kg}^{-1}\text{K}^{-1}$ for s . This has to be obtained by simplification. Once you get the answer, 429 is not in the table. Then the children will get upset. So, they may tend to redo the question thinking that they went wrong. Next, by thinking that 450 is near to 429, many children may have taken iron. Instead of dividing $900/2.1$, if $900/2$ then you will get 450. As 900 is not exactly divided by 2.1, one can think that the examiners may have been sympathetic and may have set the answer to 450. It is fair to think like that in one way.
- (2) But these arguments do not give the answer that the examiners expected. Therefore, we need to find another way. I know these will not come to your head when answering the paper. If you think of the heat loss accidentally (question does not mention to neglect the heat loss), then you will think that the real s value is less than the calculated s value. If so, you will pick up copper

with a lesser value which is near to 429. I heard there were children who selected copper without any calculations. Because we most of the time use copper containers/ copper calorimeter in such calculations and experiments. It is luck by chance.

- (3) Selecting a silver container is not a ground reality. The daughters of Saudi king put water in silver containers and put lead balls. Even Saudi king had made golden commodes for his daughters to be used in the bathroom. They may have a nice smell when answering the call of nature. Even drinking water in lead containers is not good for health as lead (Pb) is poisonous.

However, I will not see as a wrong thing because in such calculations we have been instinctively practiced to neglect the heat loss and expect/ be conscious about a minimum heat loss. So, we do not have any alternative than selecting a lower and a near value of 429 which is 385 as the answer.

43. Three right angled prisms of refractive indices n_1 , n_2 , and n_3 ($n_2 > n_1, n_3$) are arranged very close to each other on a table as shown in the figure. There are no gaps between the contact surfaces of the prisms. A ray entering through the face AB with an incident angle i , refracts at faces AB, BC, CD, and DE, and emerges from the face DE without deviation. The angles of refraction at the faces AB, BC, and CD are r_1 , r_2 , and r_3 , respectively. Which of the following expressions is incorrect?
- (1) $\sin i = n_1 \sin r_1$ (2) $n_2 \sin r_2 = n_1 \cos r_1$ (3) $\sin i = n_3 \cos r_3$
 (4) $n_2 \cos r_2 = n_3 \sin r_3$ (5) $\cos i = n_3 \cos r_3$

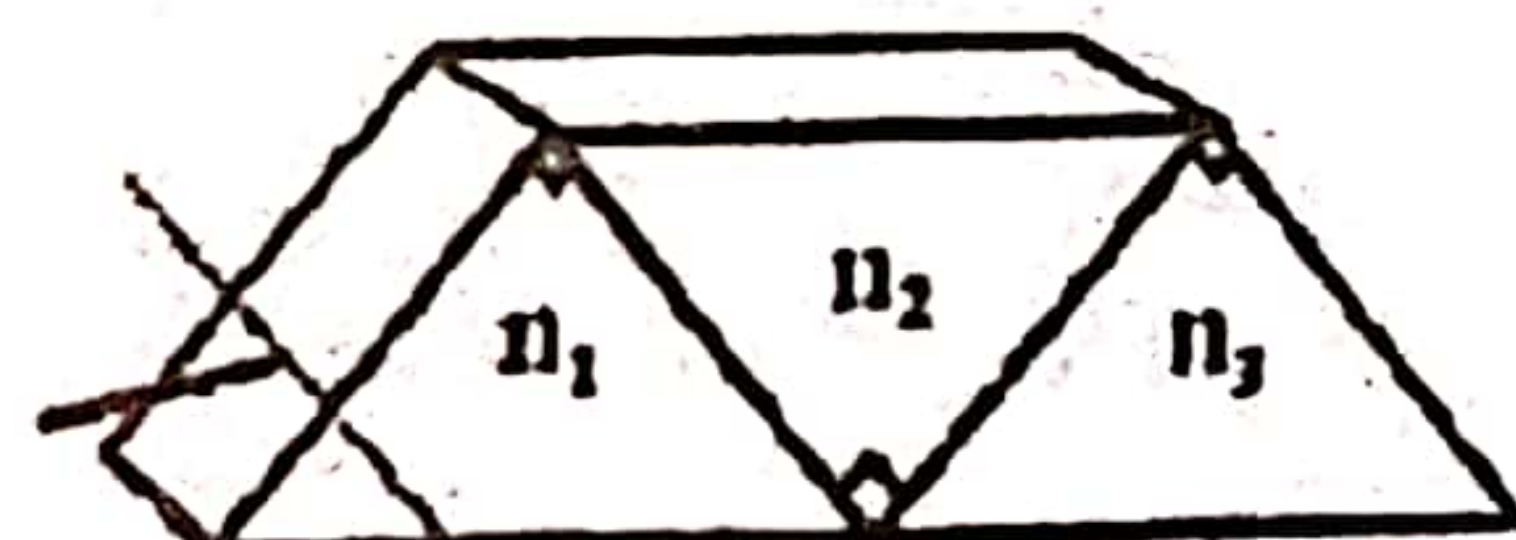


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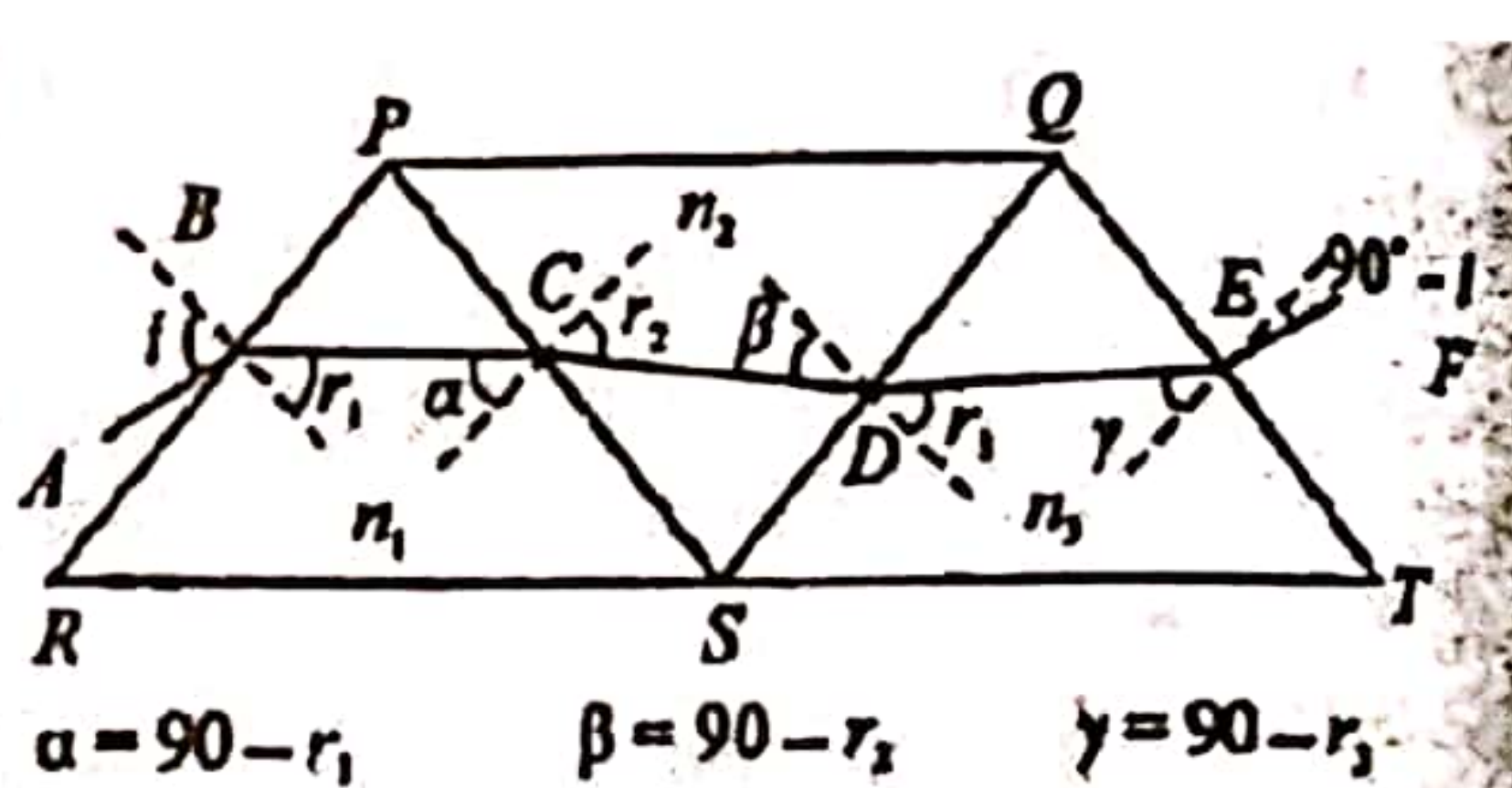
Refraction Through Prisms

Here it has shown a question taken from the Internet.

Three right angled prisms of refractive indices n_1 , n_2 and n_3 are fixed together using an optical glue as shown in the figure. A monochromatic ray passes through the prisms without suffering any deviation and come out from the last surface.



The ray diagram of the path is shown in the second figure. The figure considered about the refractive indices of the material that the prisms v Their intension was to draw the emergent ray parallel to the incide between the refractive indices has been given ($n_2 > n_1, n_3$). Therefore, e drawn is different from the given figure, we can take the relevant argur to our question, the ray path can be drawn like this way.

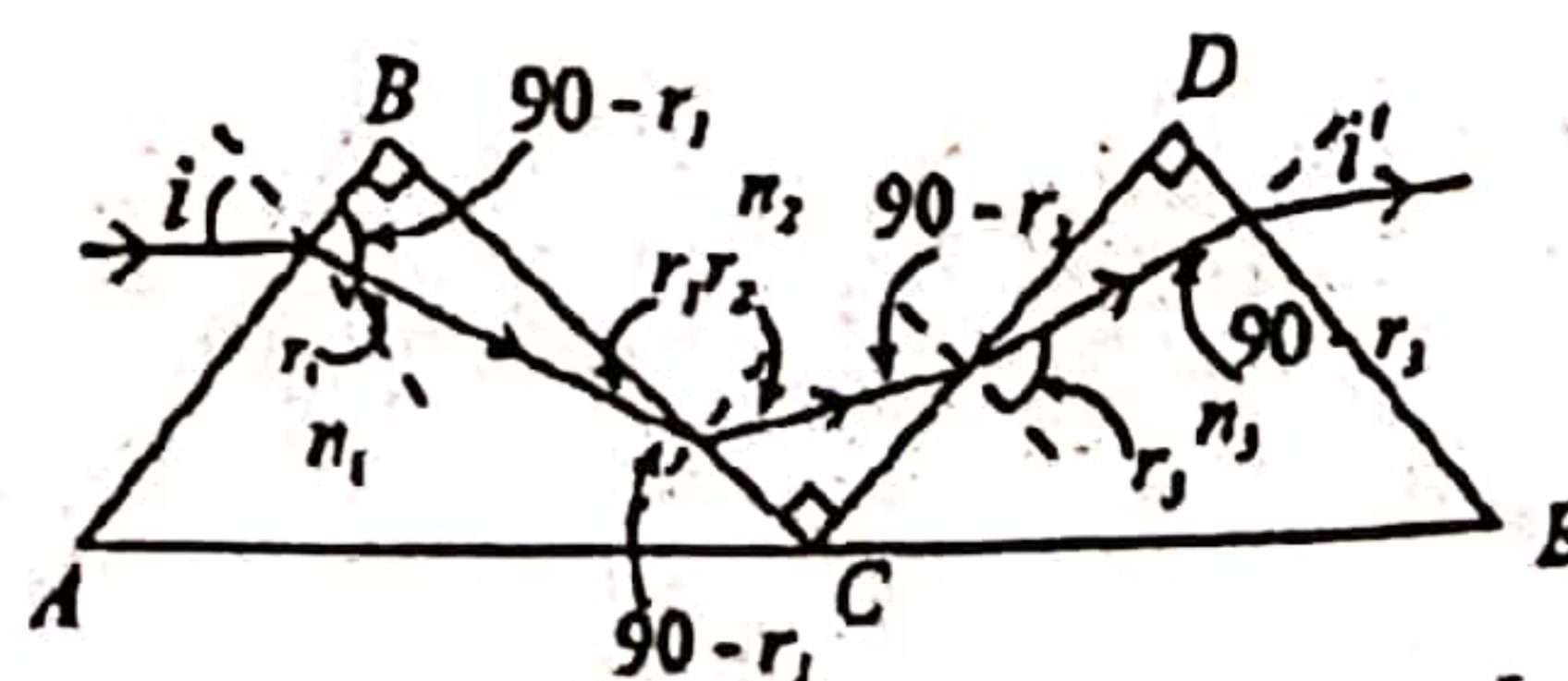


Once you draw a rough sketch like this way, it is easy to get the relations. When Snell law is applied to surface AB, $\sin i = n_1 \sin r_1$. You should see the incident ray angle of the BC surface as $(90 - r_1)$ from geometry.

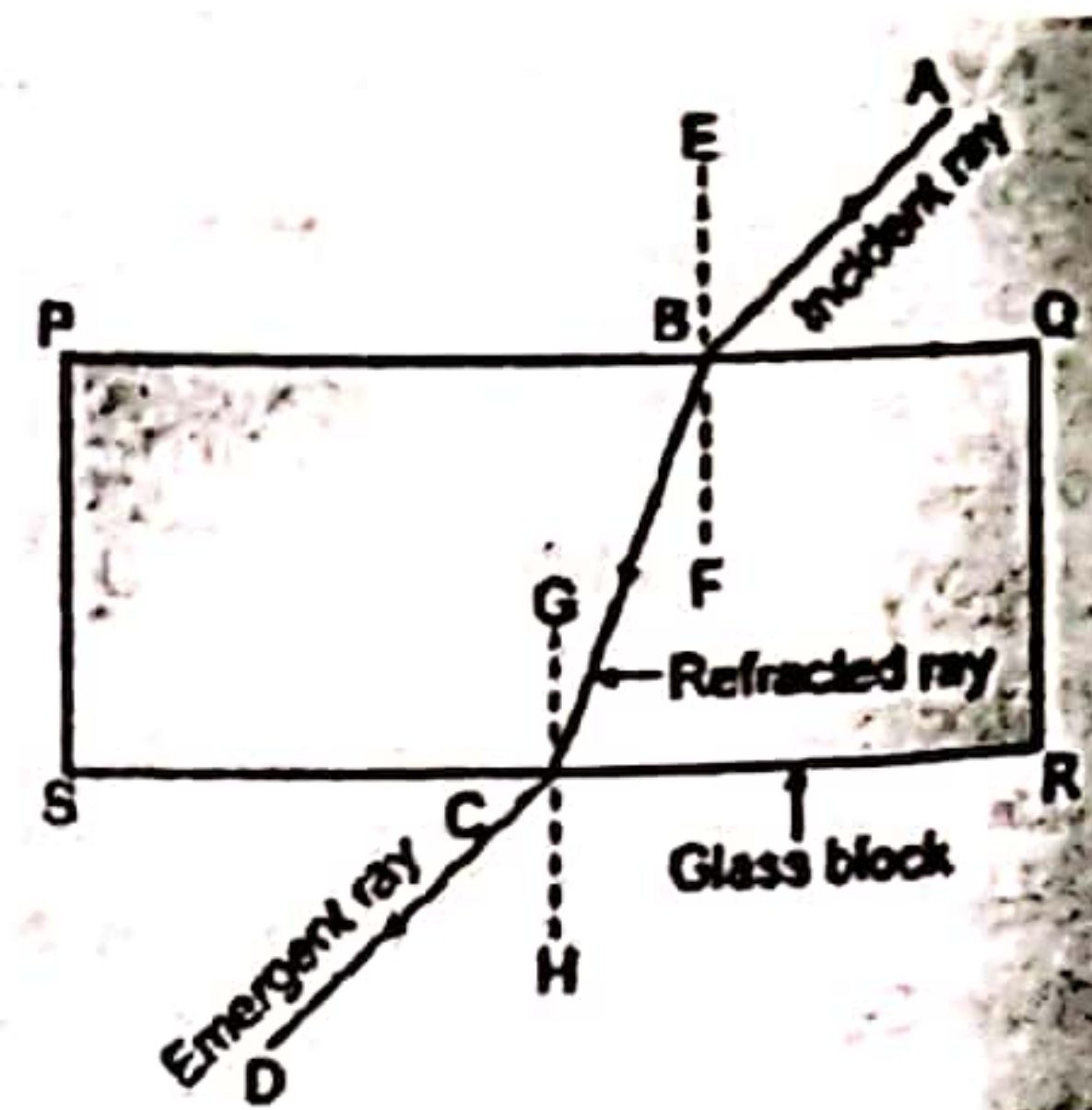
For surface BC $n_1 \sin (90 - r_1) = n_2 \sin r_2$. You should know that $\sin (90 - r_1) = \cos r_1$.

Then $n_1 \cos r_1 = n_2 \sin r_2$. Likewise for surface CD, $n_2 \sin (90 - r_2) = n_3 \sin r_3$.

Then $n_2 \cos r_2 = n_3 \sin r_3$.



Next, you need to apply Snell's law to the surface DE. The incident angle of the ray in DE surface is $(90 - r_3)$. There is no issue on that. But to decide the emergent angle from DE surface, you need to consider that the ray is not deviated. There can be a mistake here. Actually, I also got it wrong. I felt that as there is no deviation in the ray, the emergent angle also should be equal to i . You will remember a ray which will go without a deviation in a glass block with parallel sides. Look at the figure.



Here the emergent angle is the incident angle i . But at this moment, the emergent angle is $(90-i)$. A direct simple method was not seen to get this at a glance. You can get this by going across each deviation. But it is not tallying with my MCQ taste. This is a time-consuming work.

$$\text{The deviation of AB surface} = (i - r_1)$$

$$\text{The deviation of BC surface} = (90 - r_1 - r_2)$$

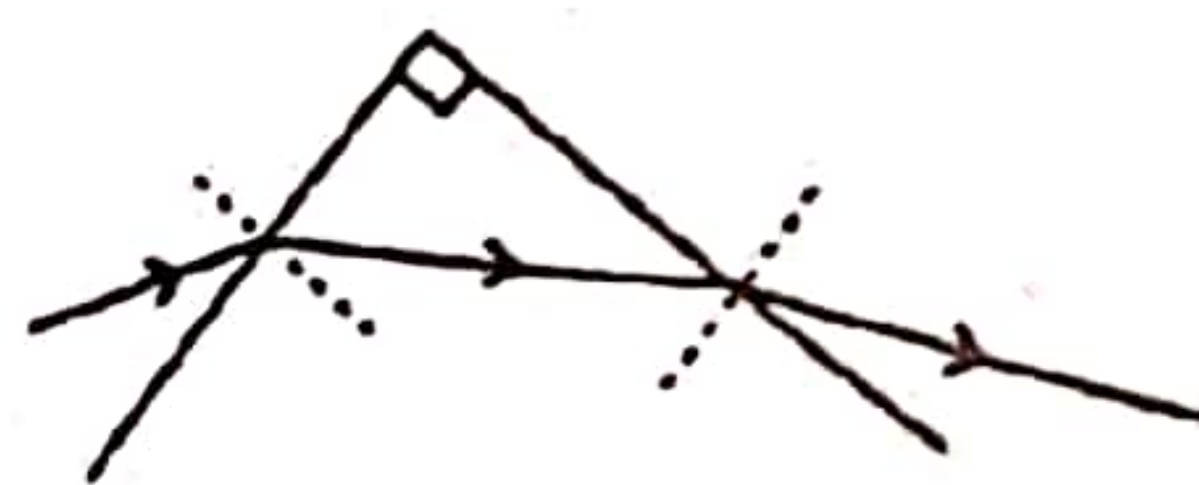
$$\text{The deviation of CD surface} = r_3 - (90 - r_2)$$

$$\text{The deviation of DE surface} = i' - (90 - r_3)$$

$$\text{If the net to be zero, then } (i - r_1) - (90 - r_1 - r_2) - [r_3 - (90 - r_2)] + i' - (90 - r_3) = 0$$

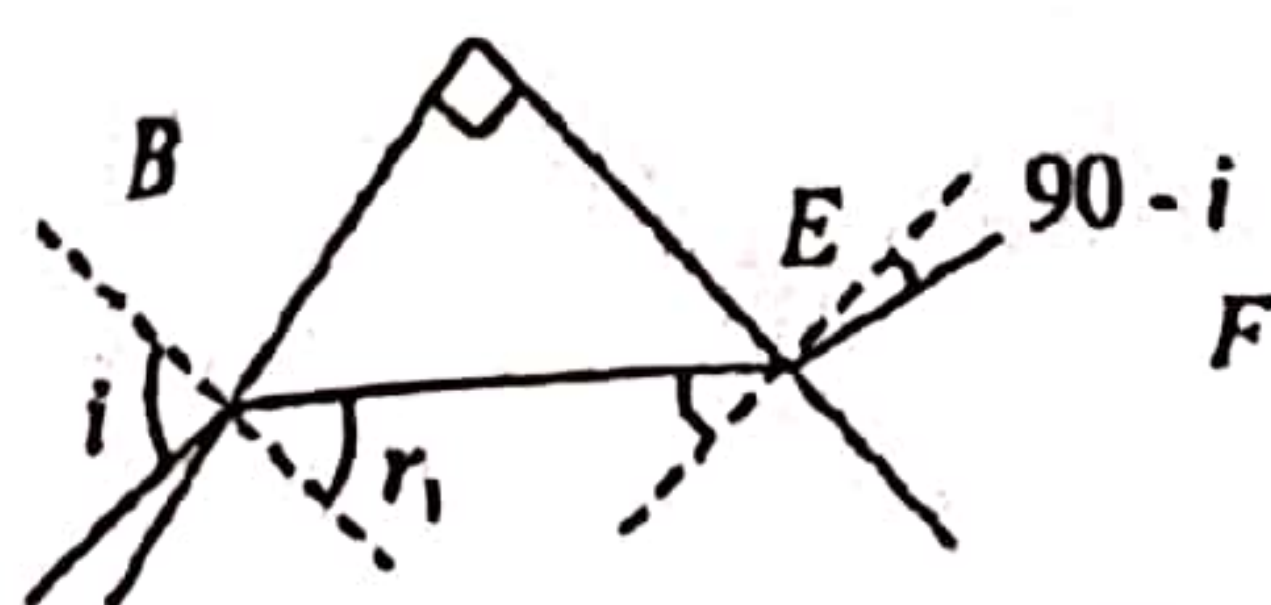
$$i' = 90 - i$$

But MCQs should not be solved like this way. As I think another way is (it is not short) to think of a big rectangular prism made from AB and DE surfaces. But the deviation cannot be made zero from a single prism refraction. Look at the figure.



The deviation from both sides is towards the same side. (↷)

But according to the other figure, if the refractive index of the emergent medium is made greater than the refractive index of the prism, then the net deviation can be made to zero.



$$\text{Now if the deviation has to be zero, then } (i - r_1) + [(90 - r_1) - i'] = 0$$

$$i - r_1 - [(90 - r_1) - i'] = 0$$

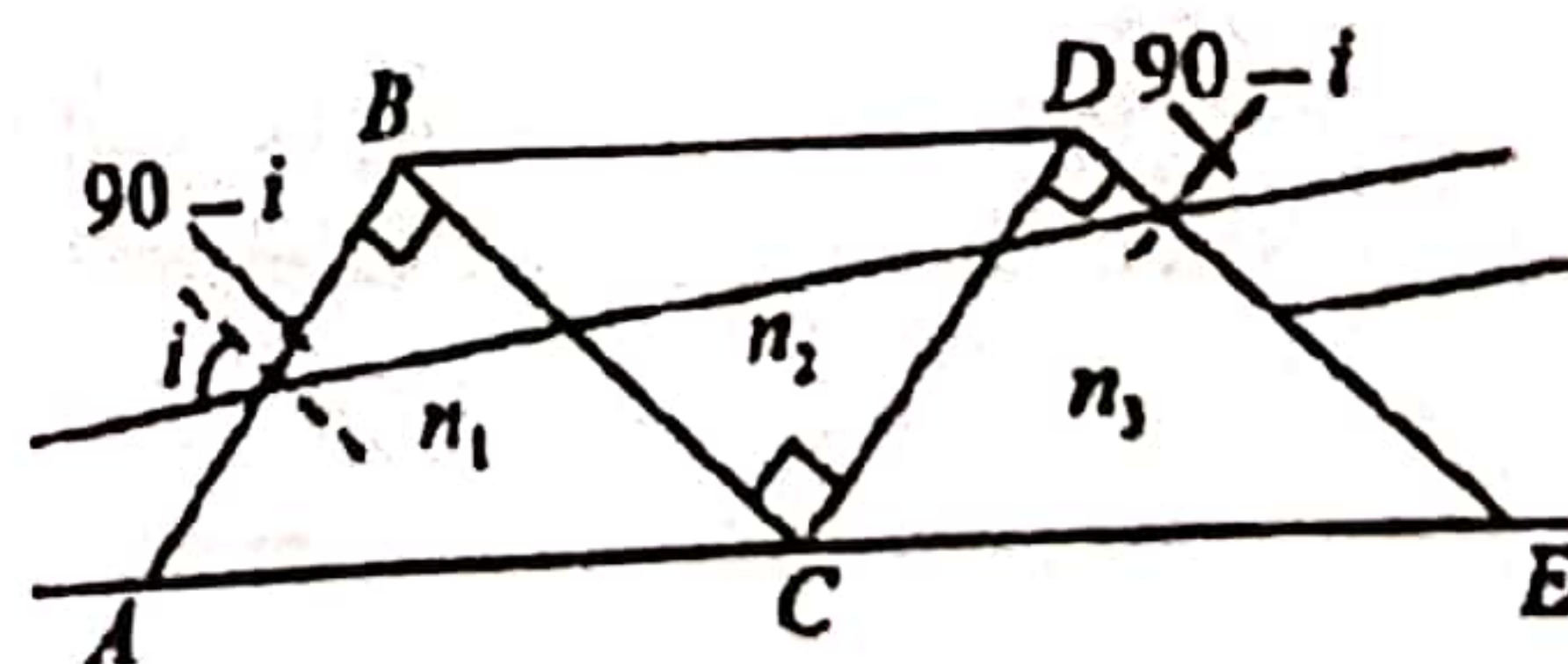
$$i' = 90 - i$$

By going across the three prisms, this work should be done.

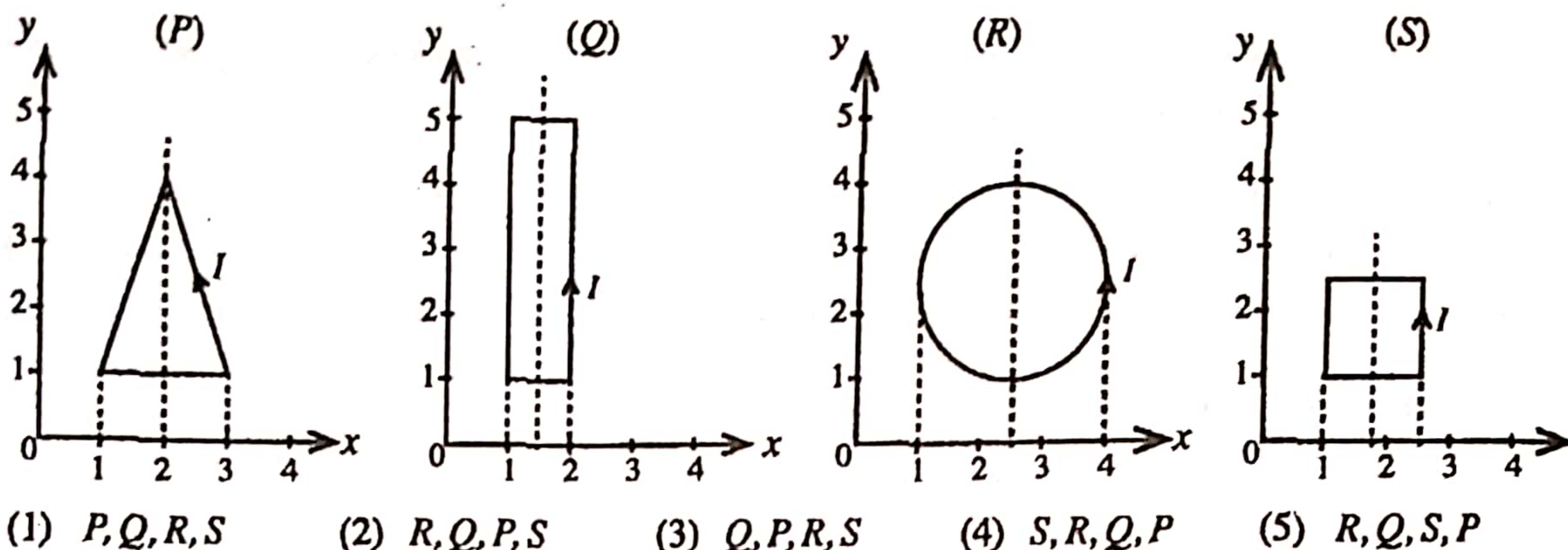
Now if we apply Snell law to the last surface, then $n_3 \sin (90 - r_3) = \sin (90 - i) \rightarrow \cos i = n_3 \cos r_3$

Time takes to solve this question. To find the emergent angle as $(90-i)$, a very simple method was suggested by a teacher. Look at the figure. The incident ray drawn in the paper lengthen it exceeding the DE surface. Then the emergent angle of the surface DE can be seen directly as $(90-i)$ because of corresponding angles.

Thank you for his suggestion.



44. Each of the single turn wire loops placed on xy plane as show'n in figures, carries the same current I . A uniform magnetic field is applied along the positive direction of the x-axis. Assume that each wire loop can rotate freely about its symmetric axis perpendicular to the magnetic field. Which choice represents the order of loops that the initial torque acting on them are in descending order?



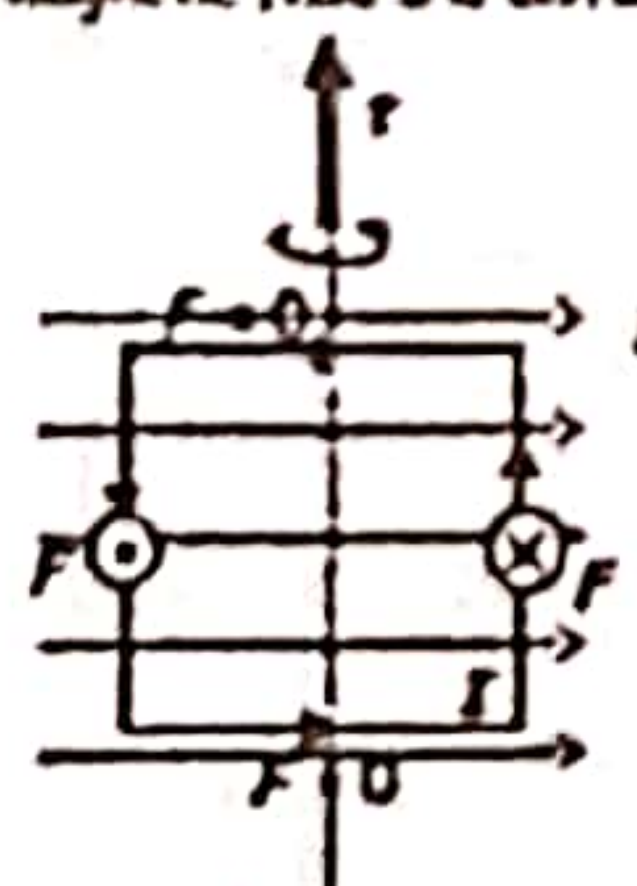
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Magnetic Field

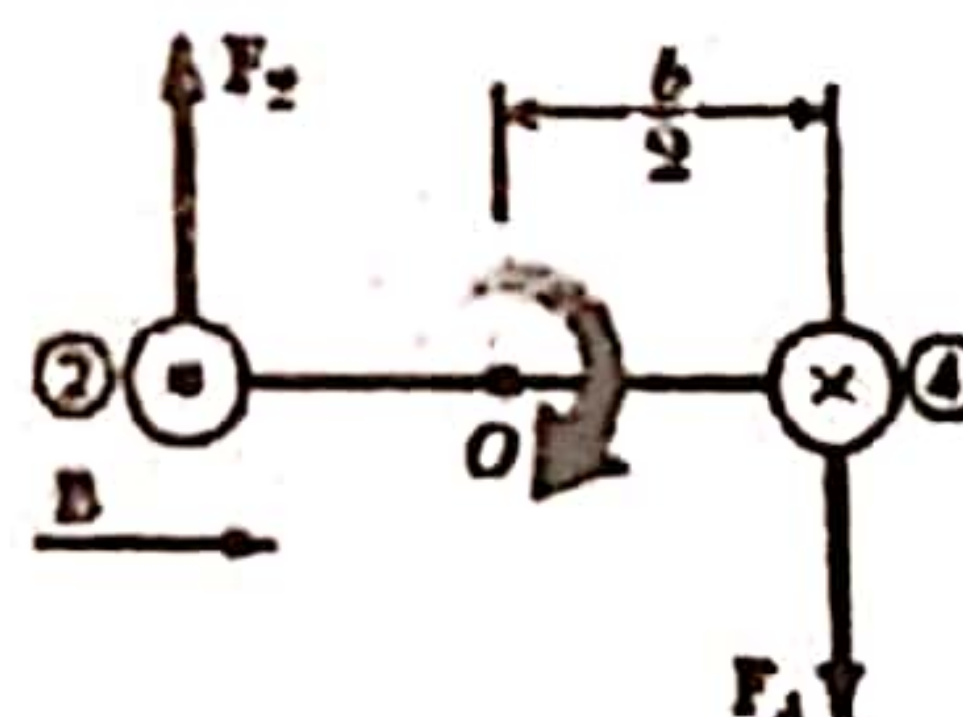
You know that the torque on a closed conducting loop which carries a current I in a magnetic field of B is $NIAB$ (N = number of turns, A = area of the loop). If there is one turn, then $N = 1$; There is no force on the horizontal wires of the loop (as I and B are parallel). There is a ILB force on the vertical left wire away from the paper and there is a ILB force towards the paper on the vertical right wire. This creates a force couple. The moments of the force couple = $F \times L = IL^2B$ (L^2 is the area of the square loop).

Torque on a square loop of current

The square loop below has side length L and carries a current I . The magnetic field B is uniform.



Net force = 0
Net torque about the center of the loop:
 $\tau = F \frac{L}{2} = I L \frac{L}{2} B$
 $= IL^2B$



You need to calculate the area of the loop. The loop with the highest area has a highest torque whereas the loop with the least area has a lowest torque. Look at the loops of our question.

Four wires each of length 2, are bent into four loops P , Q , R and S and then suspended into uniform magnetic field. Same current is passed in each loop. Which statement is correct?



- (a) Couple on loop P will be the highest
 (b) Couple on loop Q will be the highest
 (c) Couple on loop R will be the highest
 (d) Couple on loop S will be the highest

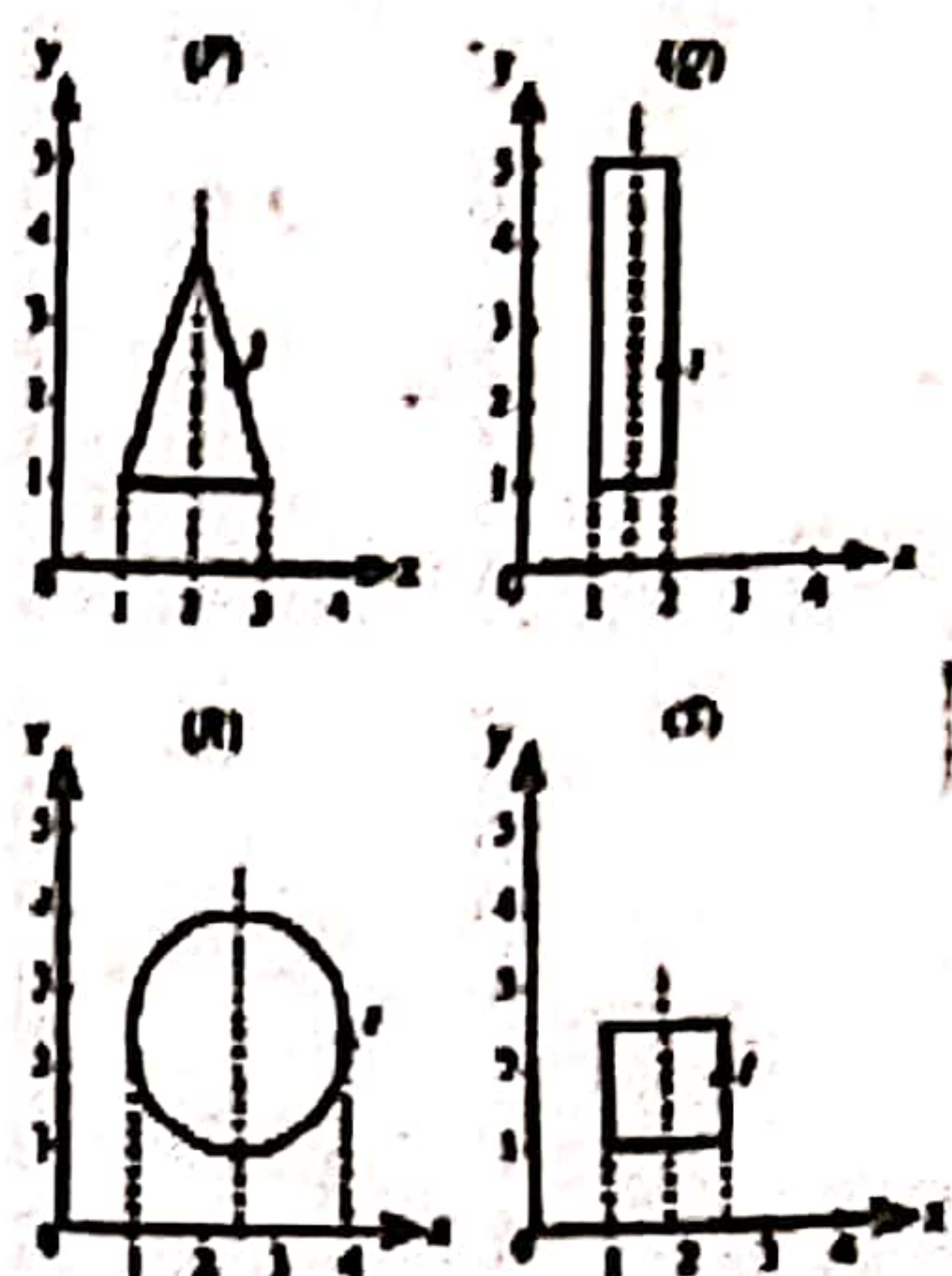
$$\text{Area of P (triangle)} = \frac{1}{2} \times 2 \times 3 = 3 \left[\frac{1}{2} \times (3-1) \times (4-1) \right]$$

$$\text{Area of Q} = 1 \times 4 = 4$$

$$\text{Area of R} = \pi r^2 = 3 \times (3/2)^2 = (3 \times 9)/4 \approx 6.8 \text{ (take } \pi = 3 \text{)}$$

$$\text{Area of S} = 1.5 \times 1.5 = 3/2 \times 3/2 = 9/4 = 2.25$$

Highest is R, then Q; next is P and finally S. \rightarrow R, Q, P, S.



45. Three cells with electromotive force (emf) E_1 , E_2 , and E_3 , and internal resistances r_1 , r_2 , and r_3 , respectively, are connected as shown in the figure. Which of the following expressions gives the potential at point P of the circuit?

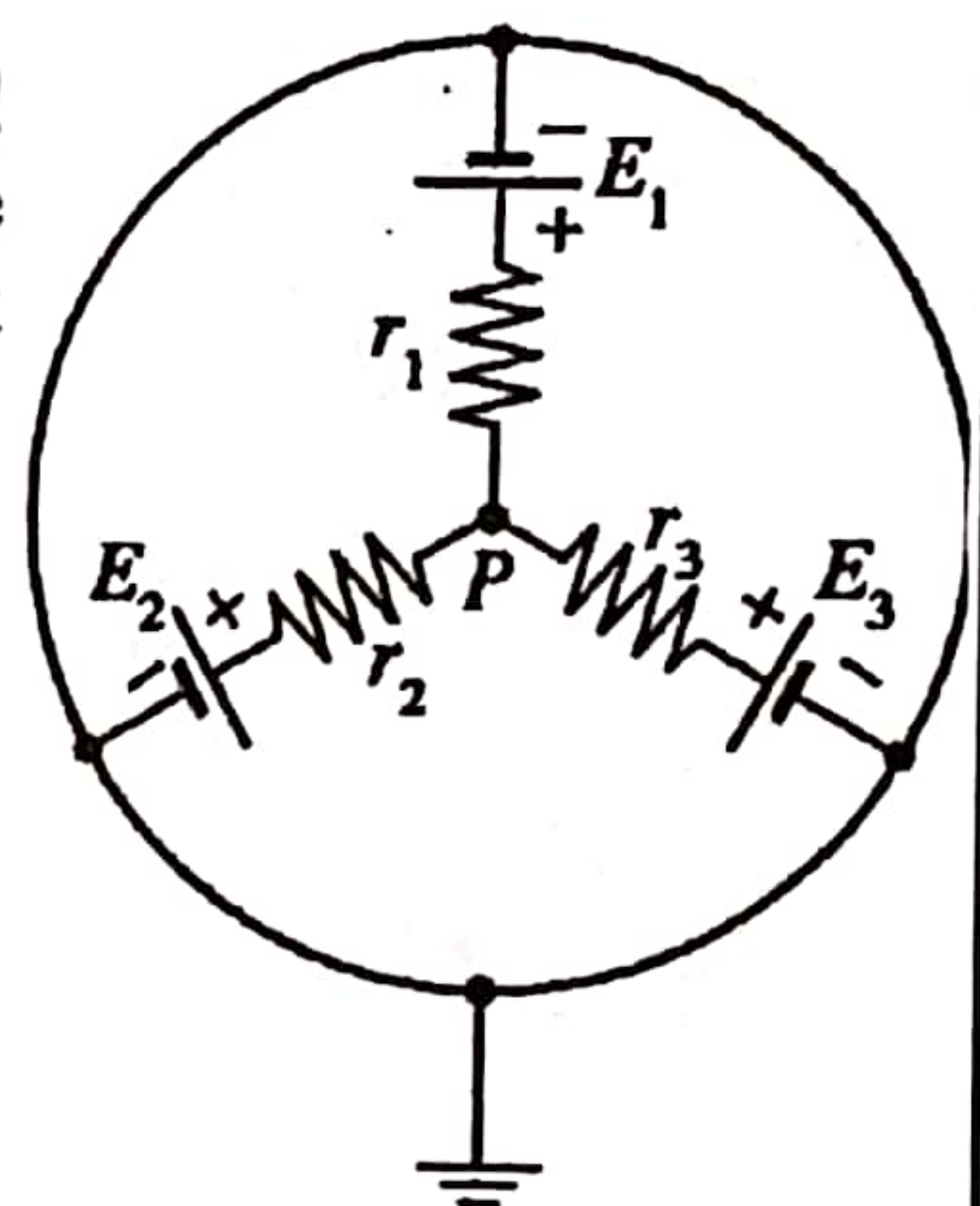
(1) $\frac{E_1 + E_2 + E_3}{3}$

(2) $\frac{E_1 E_2 E_3}{E_1 E_2 + E_2 E_3 + E_3 E_1}$

(3) $\frac{E_1 r_1^2 + E_2 r_2^2 + E_3 r_3^2}{r_1 r_2 + r_2 r_3 + r_1 r_3}$

(4) $\frac{E_1 r_2 r_3 + E_2 r_1 r_3 + E_3 r_1 r_2}{r_1 r_2 + r_2 r_3 + r_1 r_3}$

(5) $\frac{E_1 r_2 r_3 + E_2 r_1 r_3 + E_3 r_1 r_2}{r_1 r_2 r_3}$



Kirchoff's Law Combination of Cells

08

This is a parallel arrangement of three cells with different e. m. f and different internal resistances. There is no such a thing in the syllabus as I know. This arrangement is equivalent to the given circuit arrangement. When it is turned like this the body will feel cool. You feel like you know this network. Applying Kirchhoff's first law to junction P,

$$i_1 + i_2 + i_3 = 0 \quad [i_1 + i_2 = -i_3]$$

$$i_1 = \frac{E_1 - V}{r_1} \quad [E_1 - i_1 r_1 = V]$$

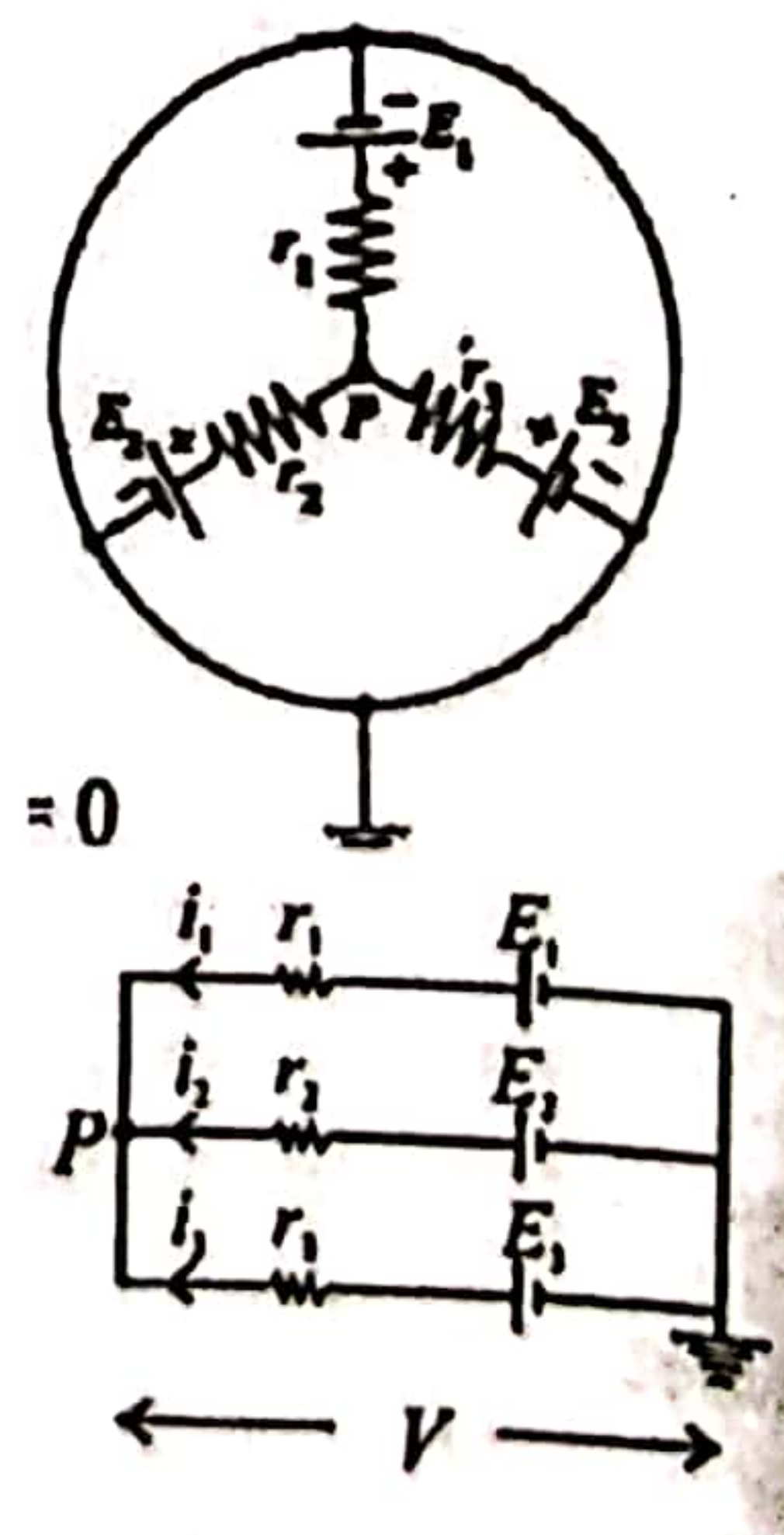
Likewise, $i_2 = \frac{E_2 - V}{r_2}$ and $i_3 = \frac{E_3 - V}{r_3}$


When they are substituted to first equation,

$$\frac{E_1 - V}{r_1} + \frac{E_2 - V}{r_2} + \frac{E_3 - V}{r_3} = 0$$

$$\frac{E_1}{r_1} + \frac{E_2}{r_2} + \frac{E_3}{r_3} = V \left[\frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3} \right]$$

$$V = \frac{E_1 r_2 r_3 + E_2 r_1 r_3 + E_3 r_1 r_2}{r_1 r_2 + r_2 r_3 + r_1 r_3}$$



There is a pattern in this expression. 

E_1 goes with $r_2 r_3$; $r_1 r_3$ goes with E_2 ; E_3 goes with $r_1 r_2$.

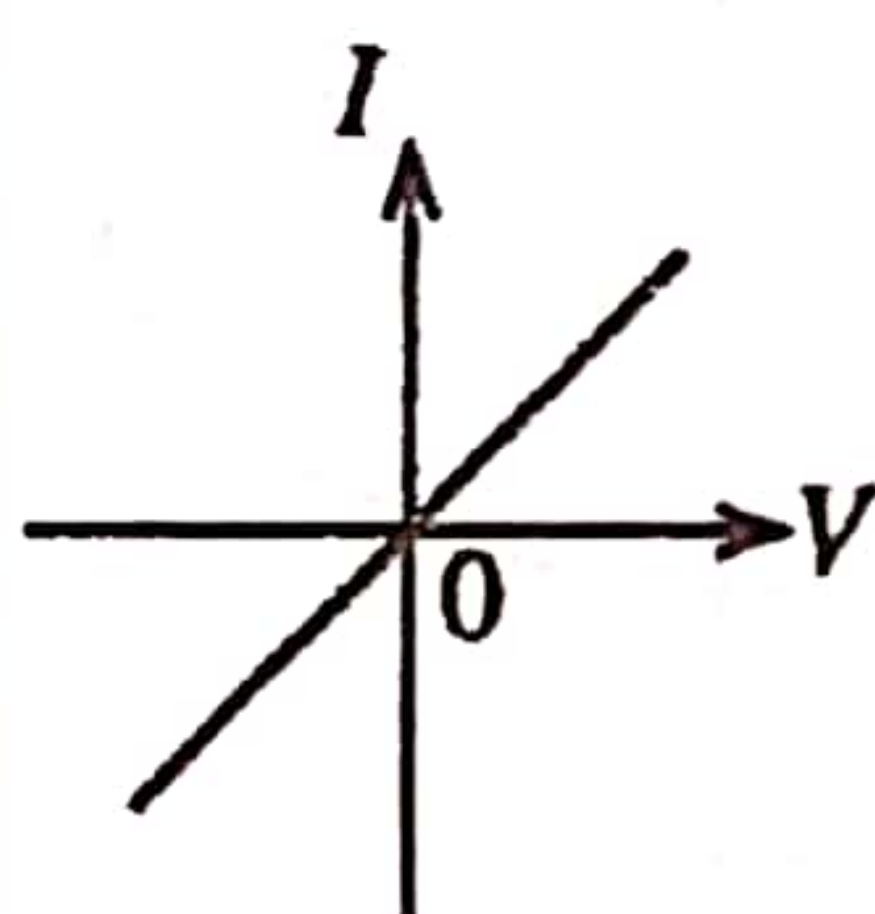
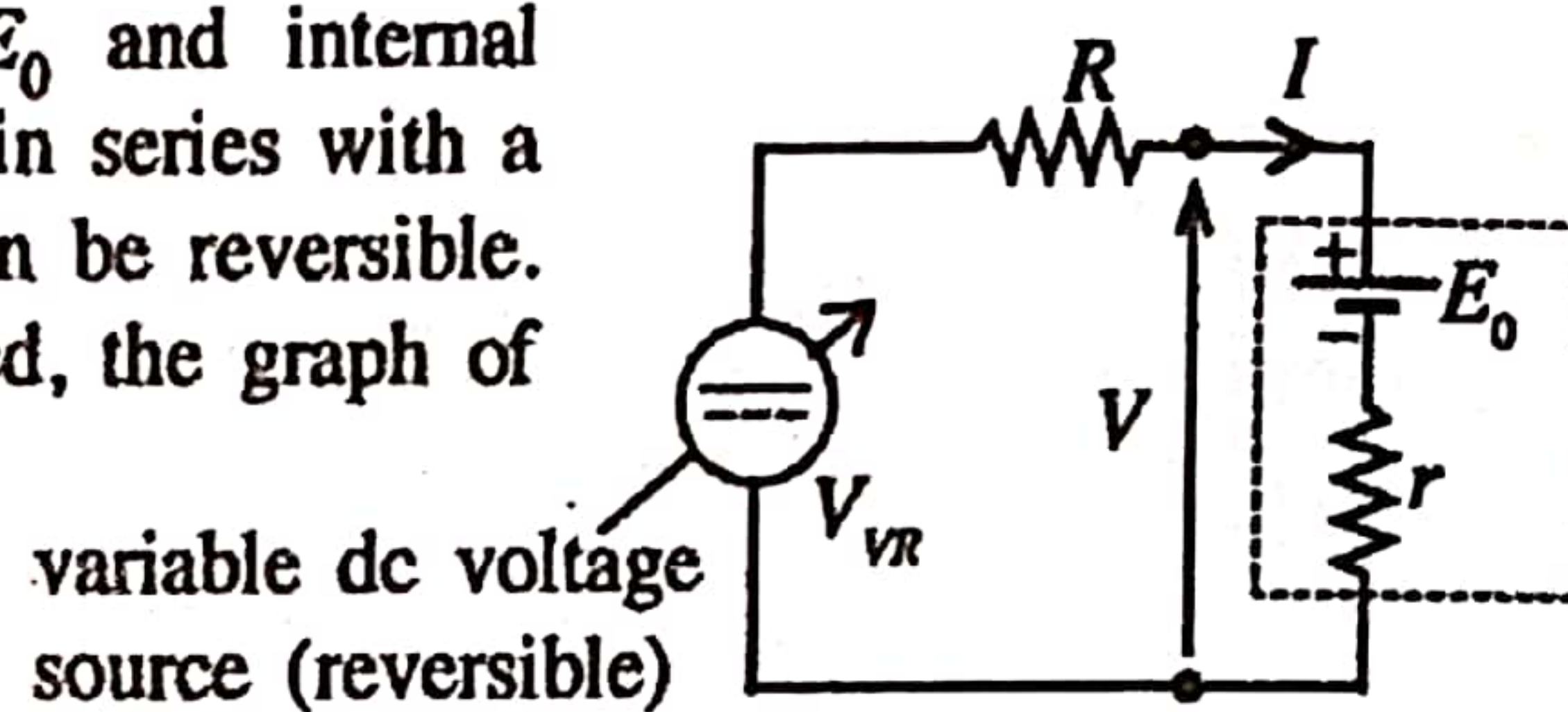
If you cannot do this, then shall I tell you a trick?

When $r_1 = 0$, $V = E_1$. When $r_1 = 0$ only the above expression gets $V = E_1$. When $r_1 = 0$,

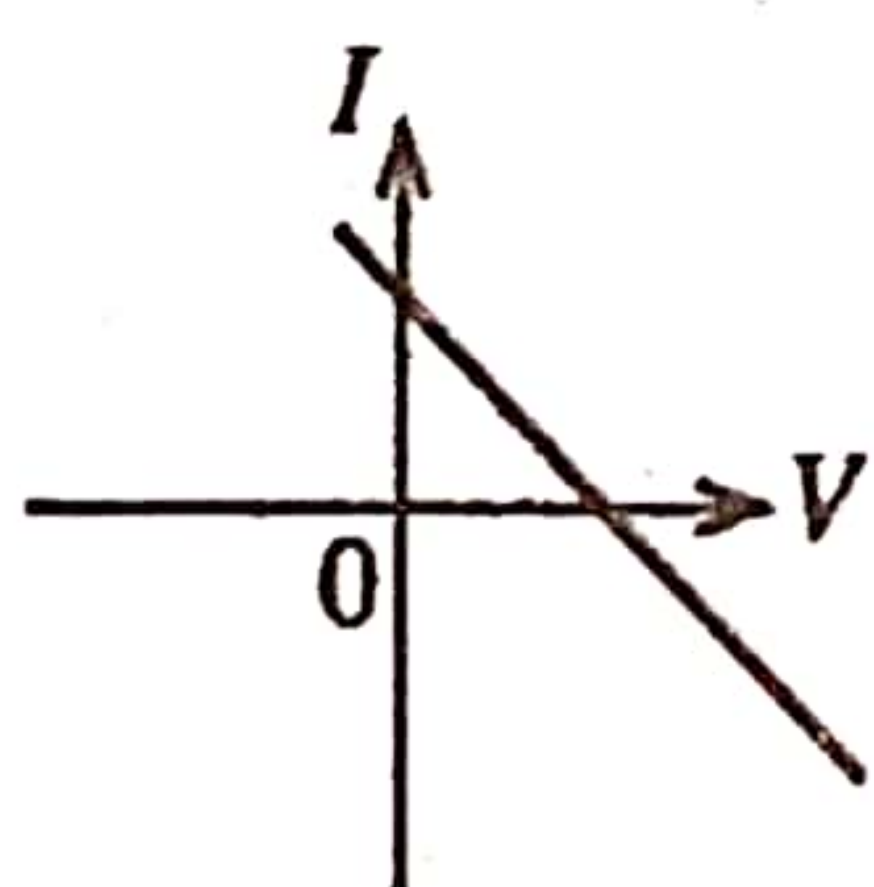
$$V = \frac{E_1 r_2 r_3}{r_2 r_3} = E_1 \text{ [} r_1 \text{ terms get zero]}.$$

Likewise, when $r_2 = 0$ $V = E_2$ and when $r_3 = 0$ then $V = E_3$. All these are satisfied by the above expression of (4). Quickly you can get the answer. But this method is dangerous where you can get many insults. My friends say that these methods unsharpened the creative ideas of children. I will let you to decide about that issue.

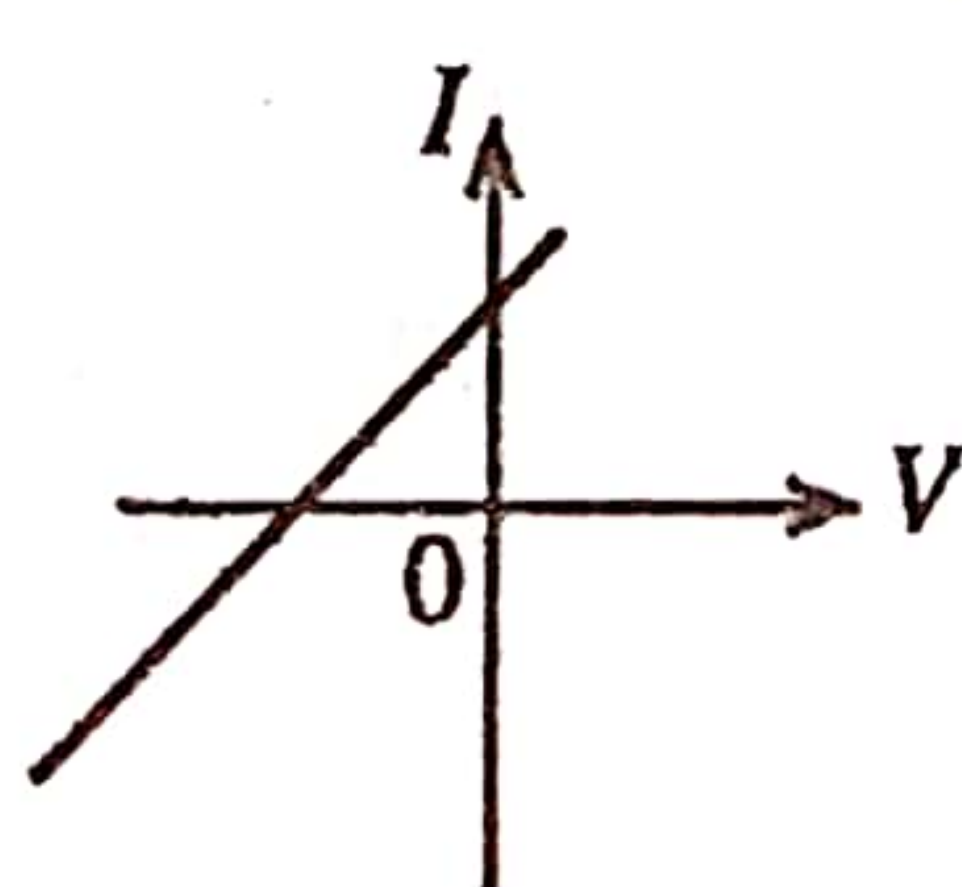
46. Consider a battery of electromotive force (emf) E_0 and internal resistance r . As shown in the figure, it is connected in series with a resistor R and a variable dc voltage source which can be reversible. When the voltage of the variable source V_{VR} is varied, the graph of I vs V is best represented by



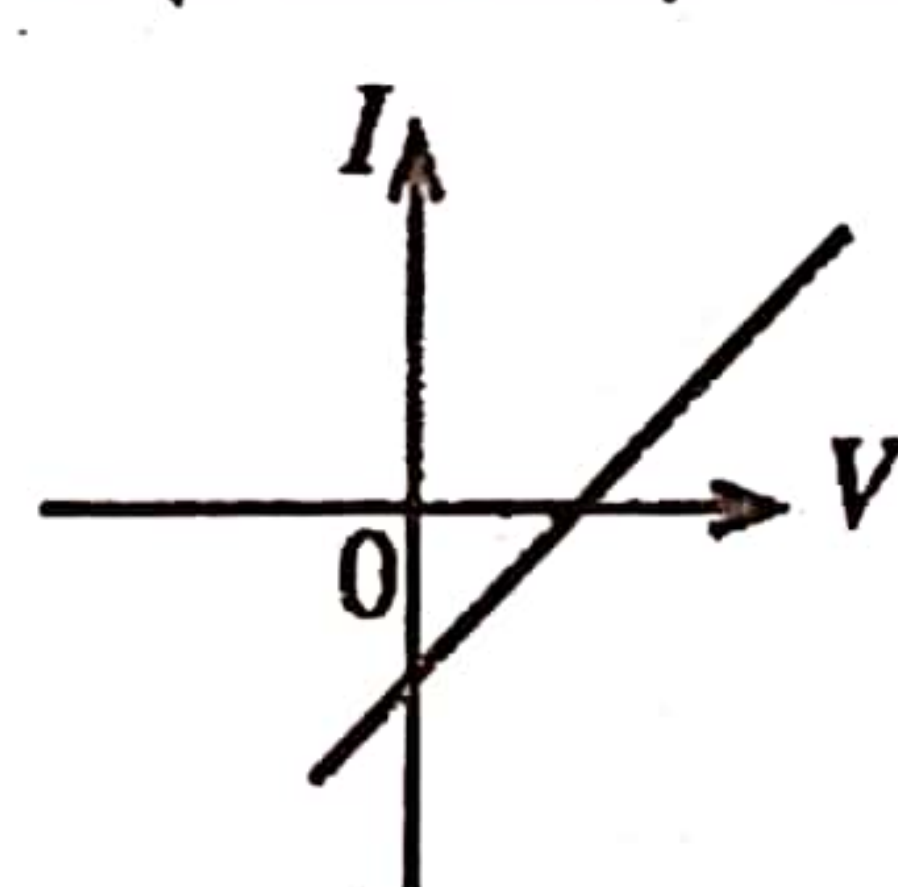
(1)



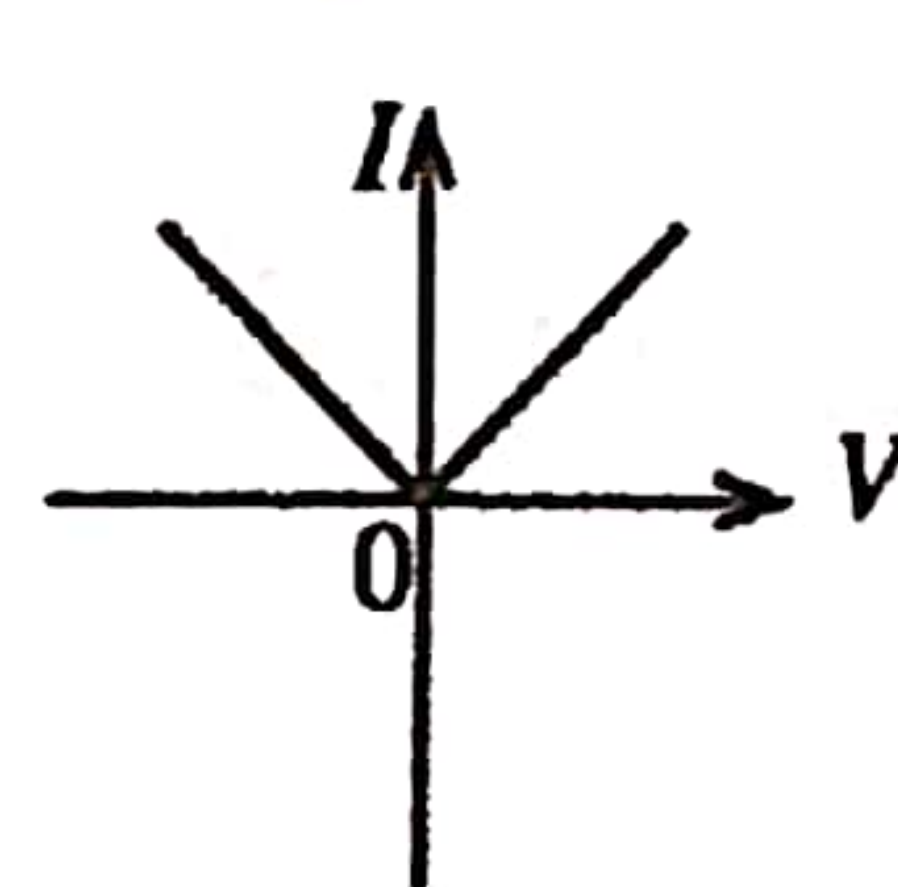
(2)



(3)



(4)

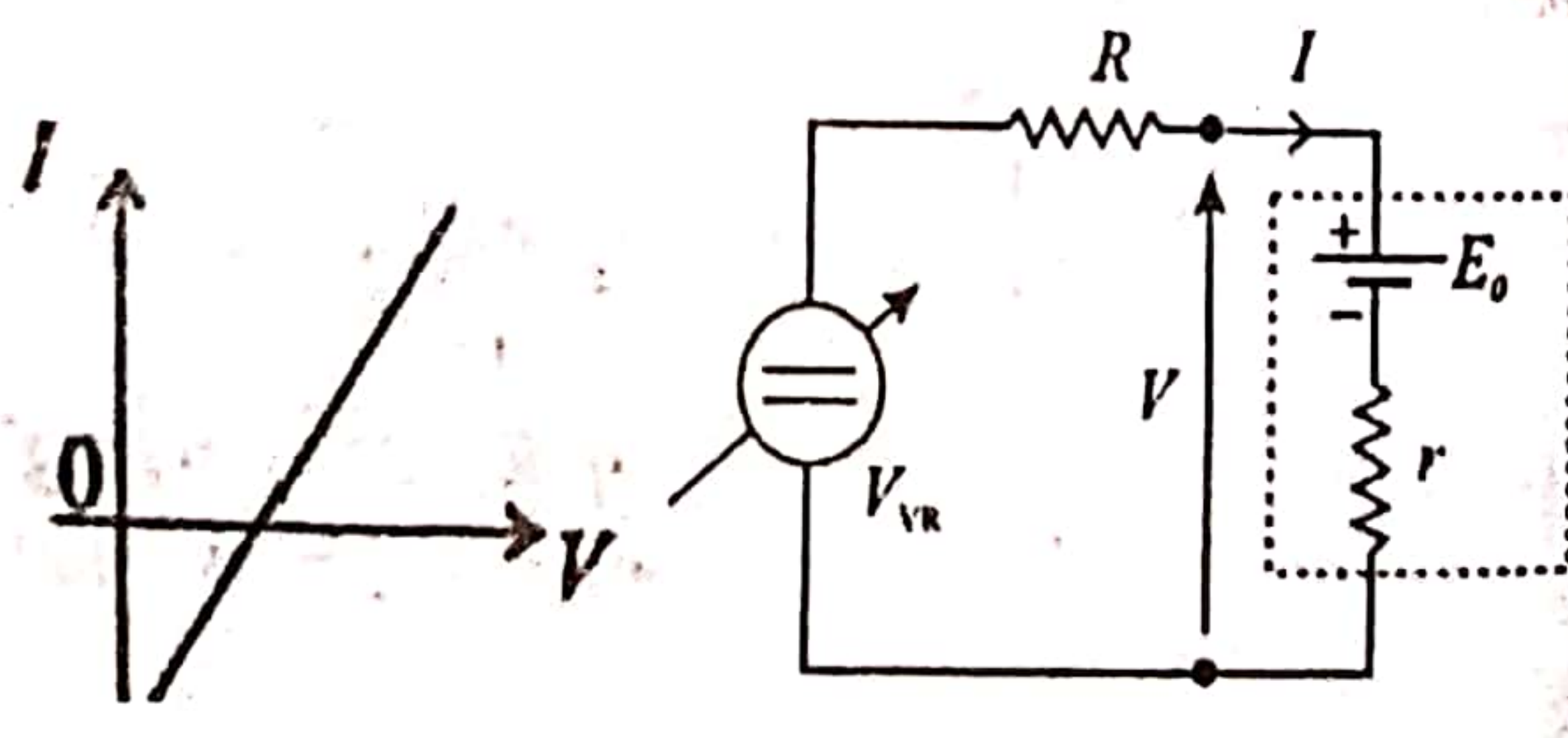


(5)

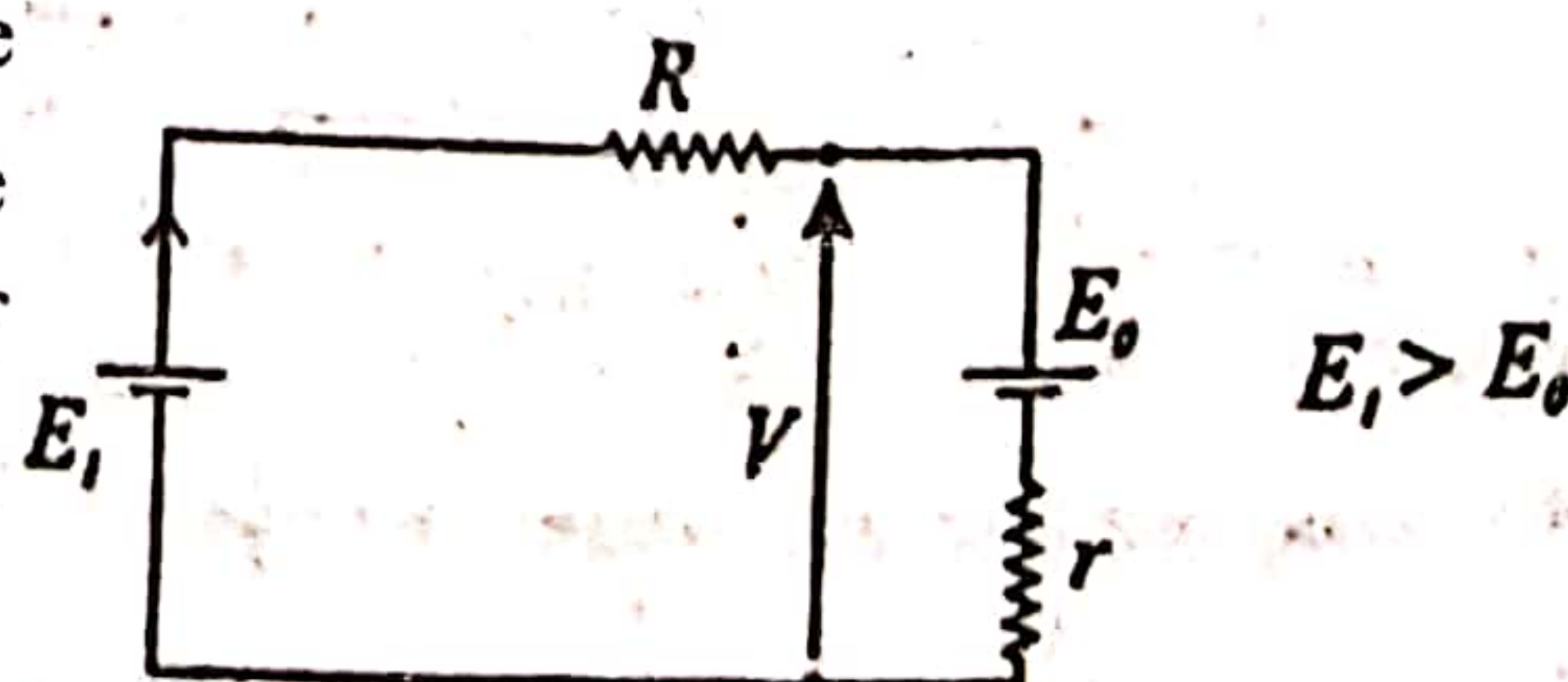
07

Korchoff's Law Combination of Celss

This is a very simple question but there is something strange too. $V = E_0 + Ir$. $Ir = V - E_0 \rightarrow I = V/r - E_0/r$. The graph of V versus I should be graph with a positive gradient and a negative intercept. If so, then the graph should be like this way. As the current is marked, we can take as $V > E_0$.

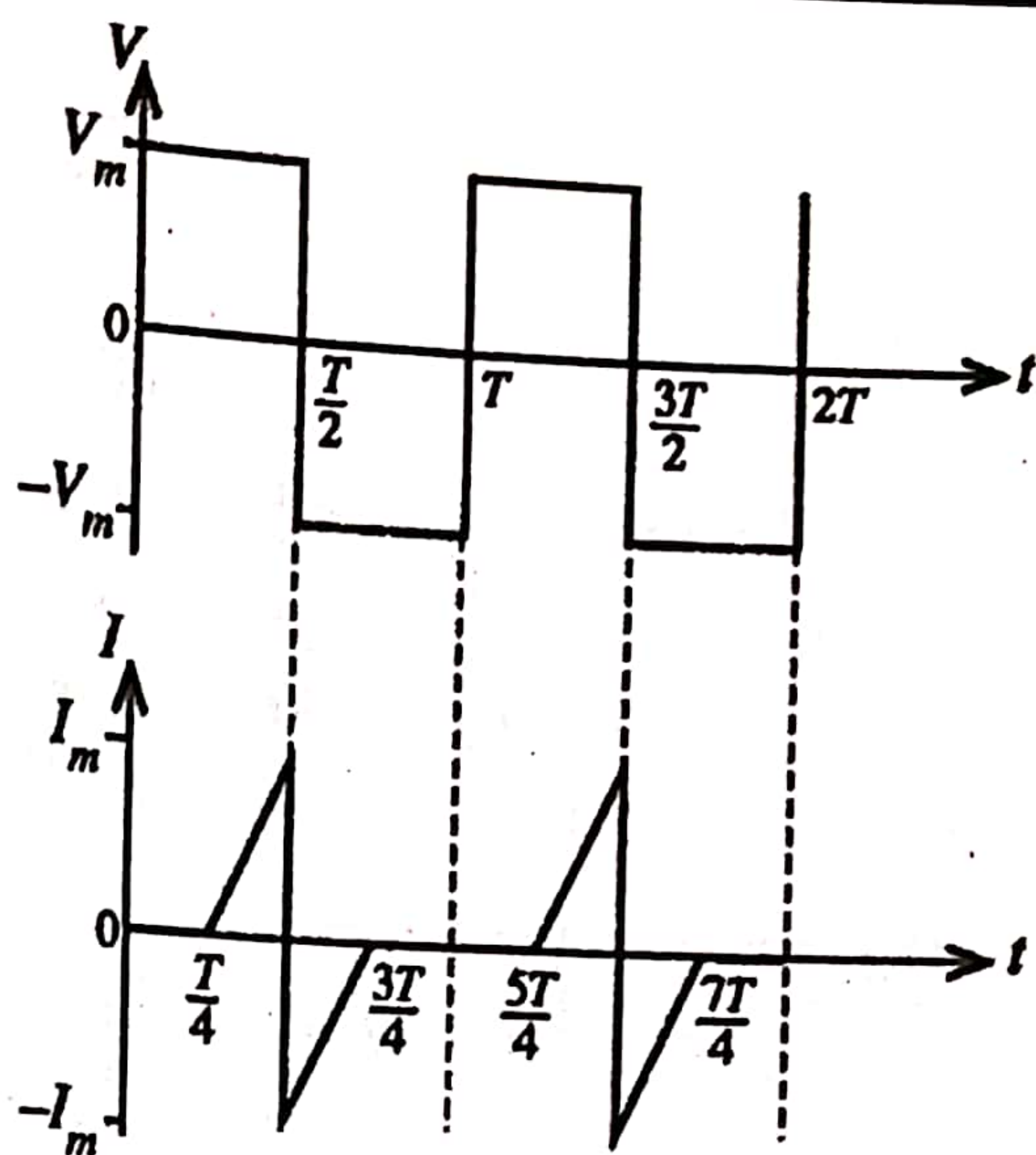
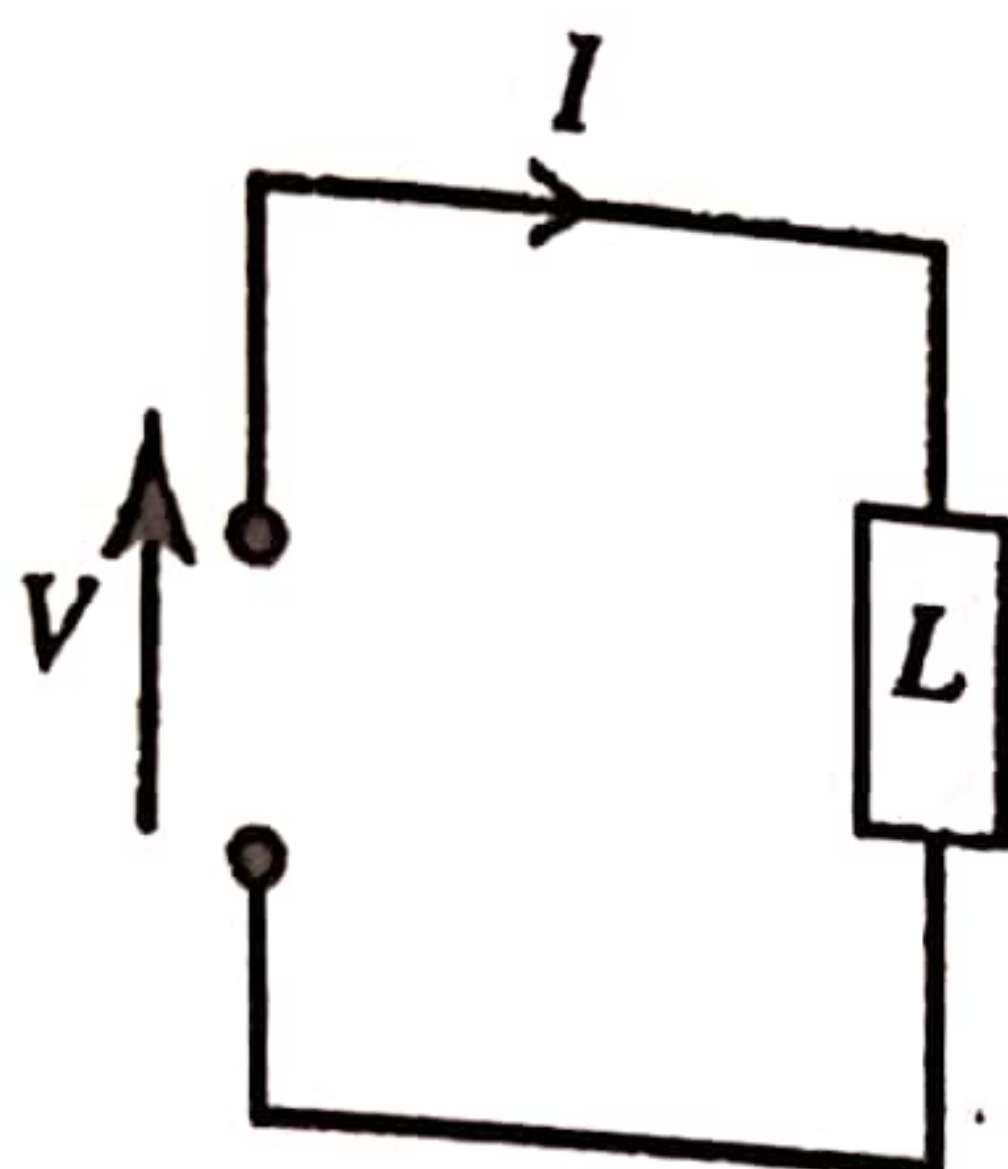


It should be considered as $V = E_0 + Ir$ not as $V = E_0 - Ir$. The variable dc voltage source sends the current into E_0 cell. If the voltage source was not there, then the current will flow out of the cell. If so, then $V = E_0 - Ir$. The current flow into the cell indicates that the voltage that is supplied from the source is stronger than E_0 . If needed, then the circuit can be shown like this way too.



$$E_1 > E_0$$

47. Consider the circuit shown in the figure. The graphs show the waveforms of the applied voltage and the current through the load L .



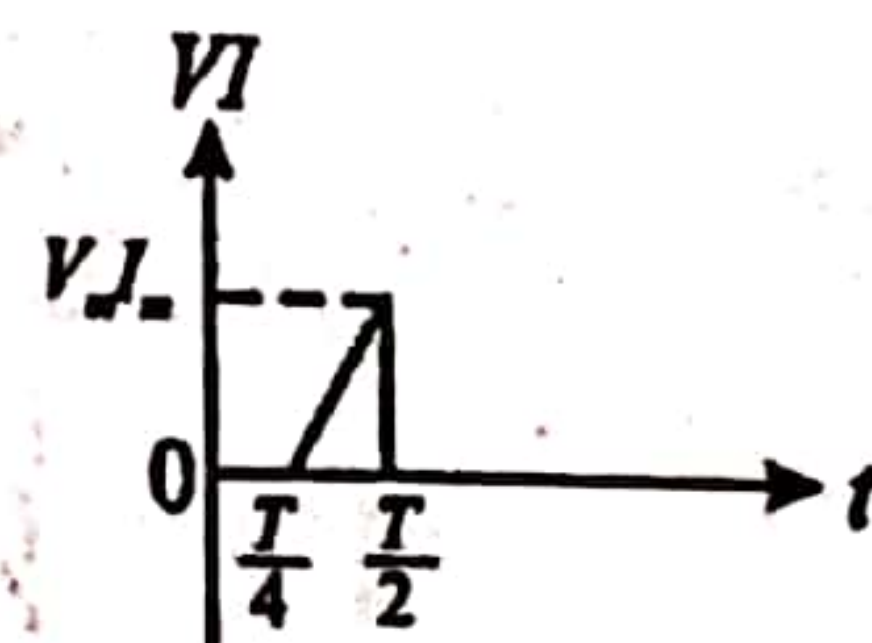
The average power dissipation of the load is

- (1) 0 (2) $\frac{V_m I_m}{4}$ (3) $\frac{V_m}{\sqrt{2}} \frac{I_m}{\sqrt{2}}$ (4) $V_m I_m$ (5) $2V_m I_m$

Heating Effect of Electric Current

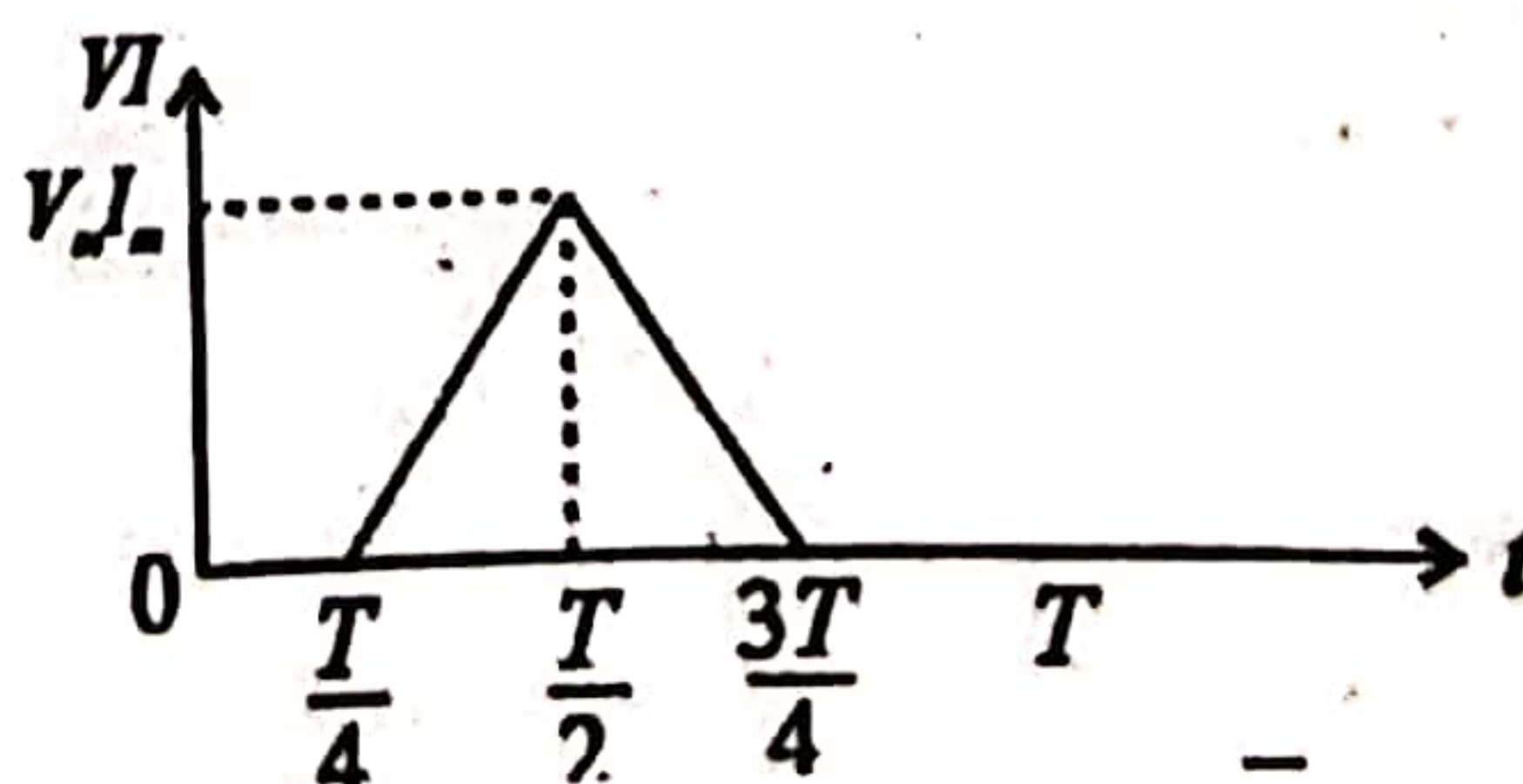
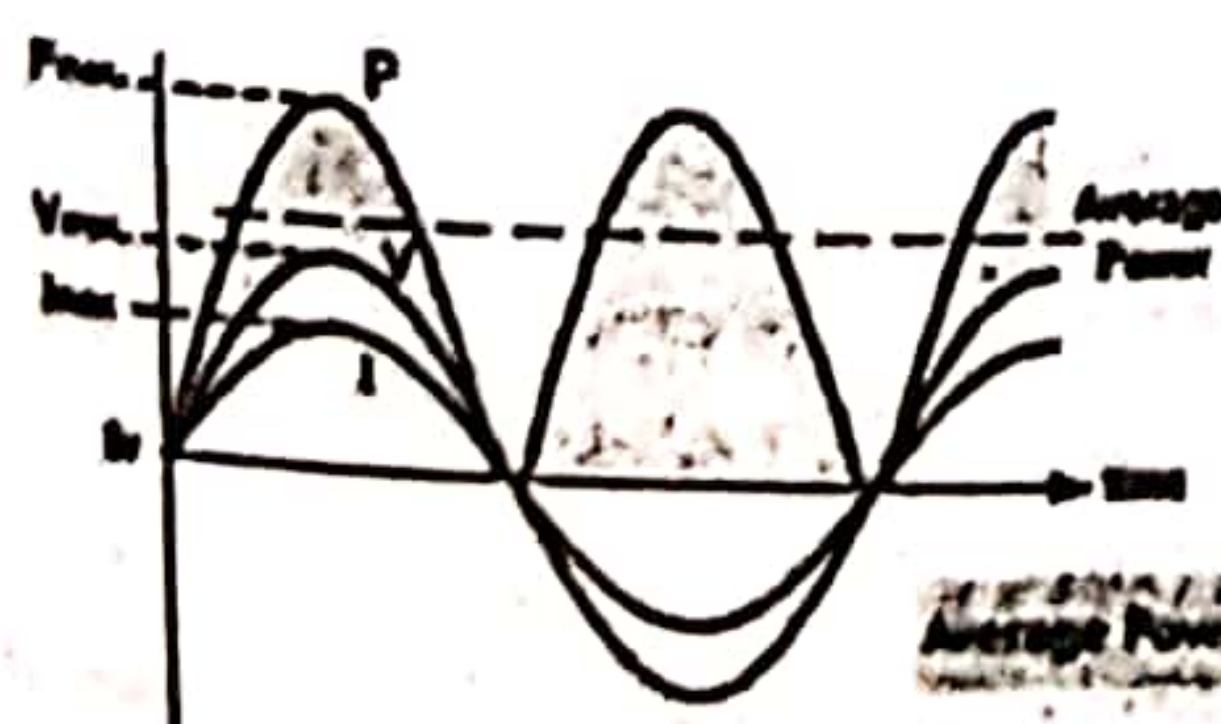
08

In a direct current circuit, the power generation is given by VI and as V and I are constant the power generation is also remains unchanged. But if V and I is changed with time, then at a certain t moment the power generation is obtained by $V(t) I(t)$. As this value varies with time, there is no usage from finding power generation for a certain moment when there are alternating currents and voltages. Therefore, calculation of the mean power generation for a period is the normal tradition. If we express this mathematically, then it will be $\bar{P} = \frac{1}{T} \int_0^T V(t) I(t) dt$. Do not get scared to this expression. What is meant by this is that finding the area from the multiplication of V and I for a total of T time (one period) and it is divided by T to get the mean. $V(t) I(t) dt$ means the power generation for a small dt time period. When this is added for a period, you will get the total energy generation during that period (in J). When this value is divided by T , you will get the mean power (in W).



Now multiply the corresponding values of V and I in a period. From $t=0$ to $t=T/4$, I is zero. Therefore, even though there is a value for V , during this time frame the multiple of VI is zero. From $t=T/4$ to $t=T/2$ multiply the constant V value with the linearly varying I . When I value comes to the top of the triangle, the VI value is $V_m I_m$. Look whether you get this variation when VI is multiplied.

After $t=T/2$, both V and I are negative. But VI multiple is positive. Now the summation of VI multiples will be the area of the triangle. It is $T/4 \times V_m I_m$ ($T/4$ means the half of the base). Therefore, mean power = $V_m I_m / 4$. $T/T = V_m I_m / 4$.



From 0 to $T/4$ and from $3T/4$ to T we are not going to divide from $T/2$ to find mean power as VI multiple is zero. It is ok that some parts are zero. The mean is found for the total period. As I remember, it is the first time such a question was given for A/L. Finding these things are not in the syllabus is my feeling. I may be wrong. Normally in A/L level we consider sinusoidal (or cosine) variations. The figure has shown the alternative voltage and alternative current variations across a resistor.

(Bio students please do not look into this.) It is clearly seen that the power generation (P) is obtained from VI multiple. If we apply mathematics, then V and I can be expressed as $V = V_m \sin \omega t$ and $I = I_m \sin \omega t$. Mean power generation (can be expressed as, $\bar{P} = \frac{1}{T} \int_0^T V_m I_m \sin^2(\omega t) dt = \frac{V_m I_m}{T} \int_0^T \sin^2(\omega t) dt$).

The value of this integral can be obtained as $T/2$. Mathematics students can try and see.

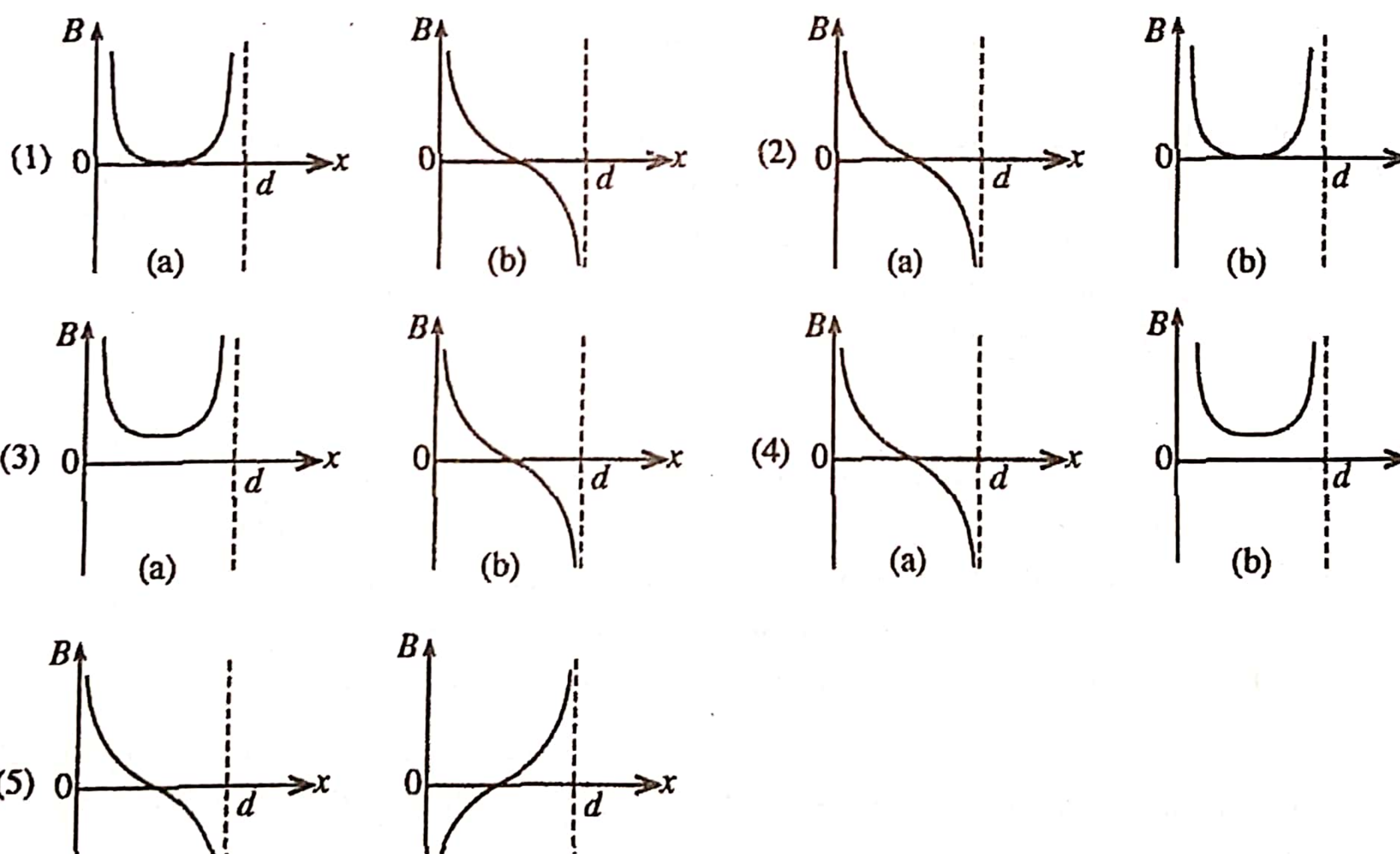
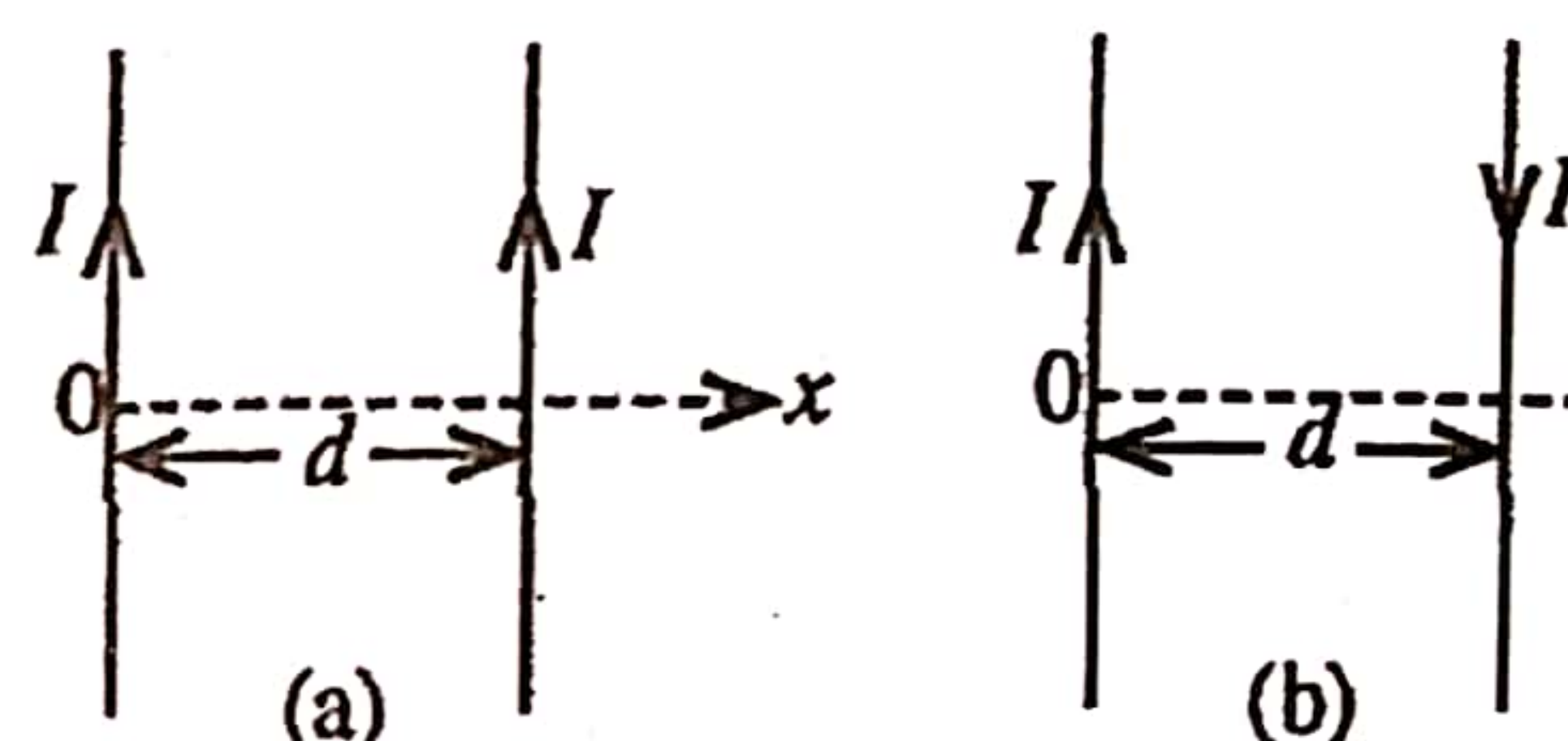
$$\bar{P} = \frac{V_m I_m}{2} = \frac{V_m}{\sqrt{2}} \frac{I_m}{\sqrt{2}} = V_{rms} I_{rms}$$

48.

Two long, straight, and parallel wires are placed in free space. Consider the following two cases as shown in the figures.

- (a) Wires carry the same current I in the same direction.
- (b) Wires carry the same current I in opposite directions.

Consider the direction of the magnetic flux density into the paper as positive. Which pair of graphs best represents, the variation of the magnetic flux density B between the two wires?

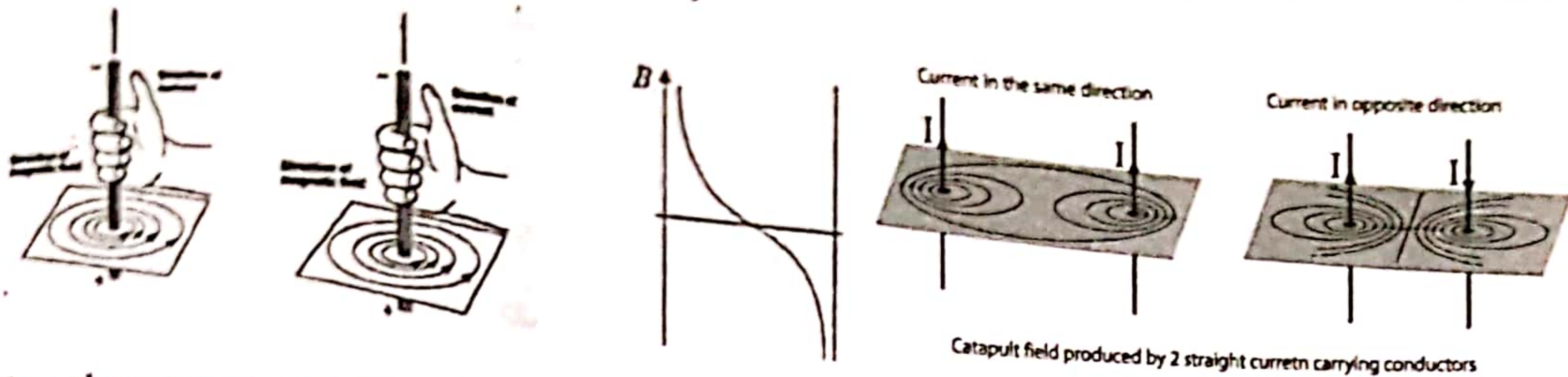


07

Magnetic effect of electric currents

A similar question was given as the 34th question in paper 2010. The figure has shown how the magnetic flux lines are placed when the currents are flowing to the same direction and different direction. You know these things very well. When the current flows to the same direction, then the magnetic flux density in the middle of wires should be zero (the null point). In between the two wires, B near to the left wire should be into the paper and B near to the right wire in between the wires should be out of the paper. Out of this, the variation of B

between the two wires should be like this way.



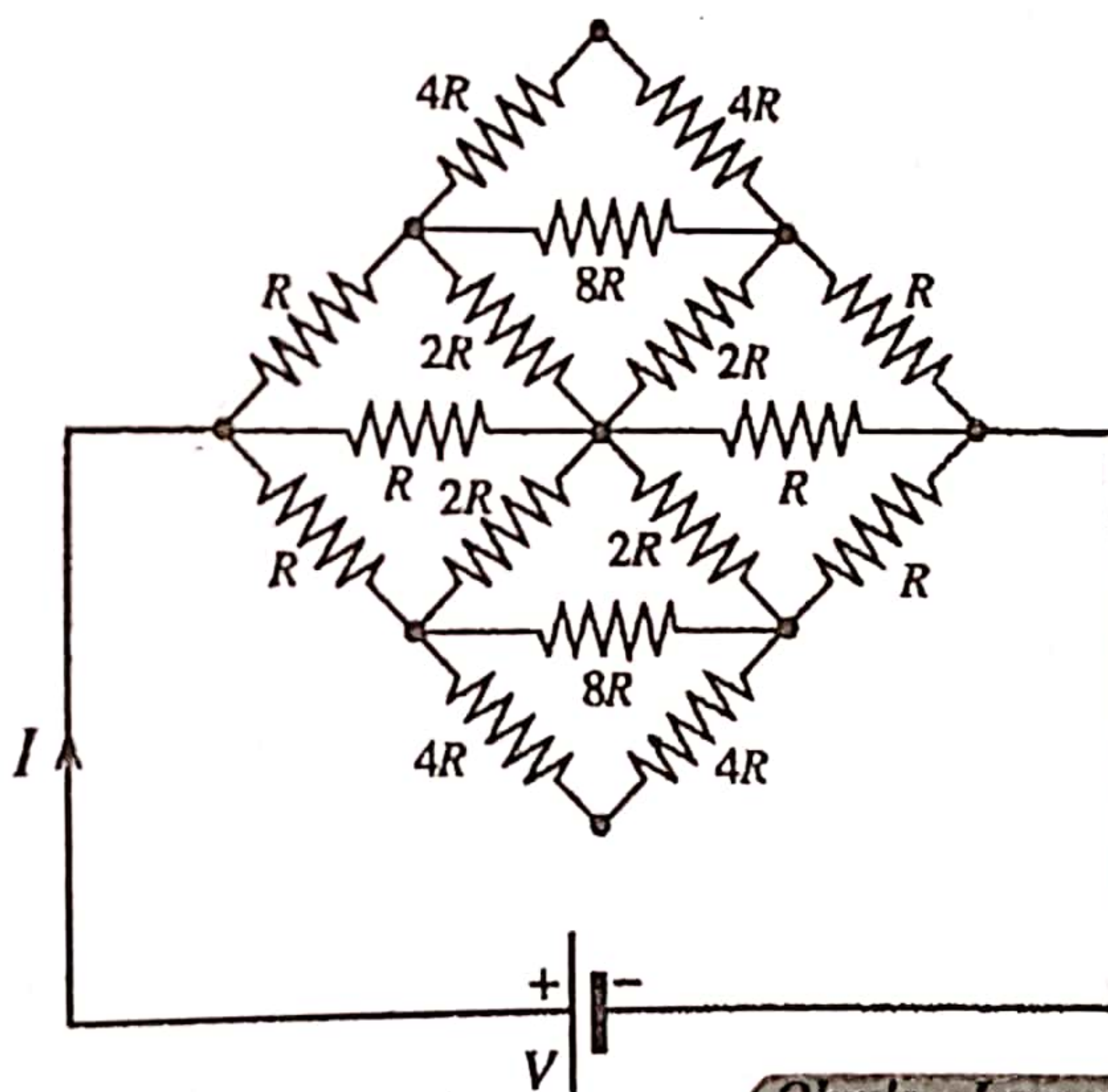
When the currents are going in the opposite direction, B in between the wires are always directed towards the paper. The middle of the wire cannot be a zero. Therefore, that variation should be like this way.



When the currents flow to the same direction, B should be zero in the middle of the wires. When the currents flow to the opposite direction, B cannot be a zero in the middle of the wires. Even if you only consider these two facts, then you will get (4) as the correct answer.

49. What is the current through the battery of the circuit shown in the figure?

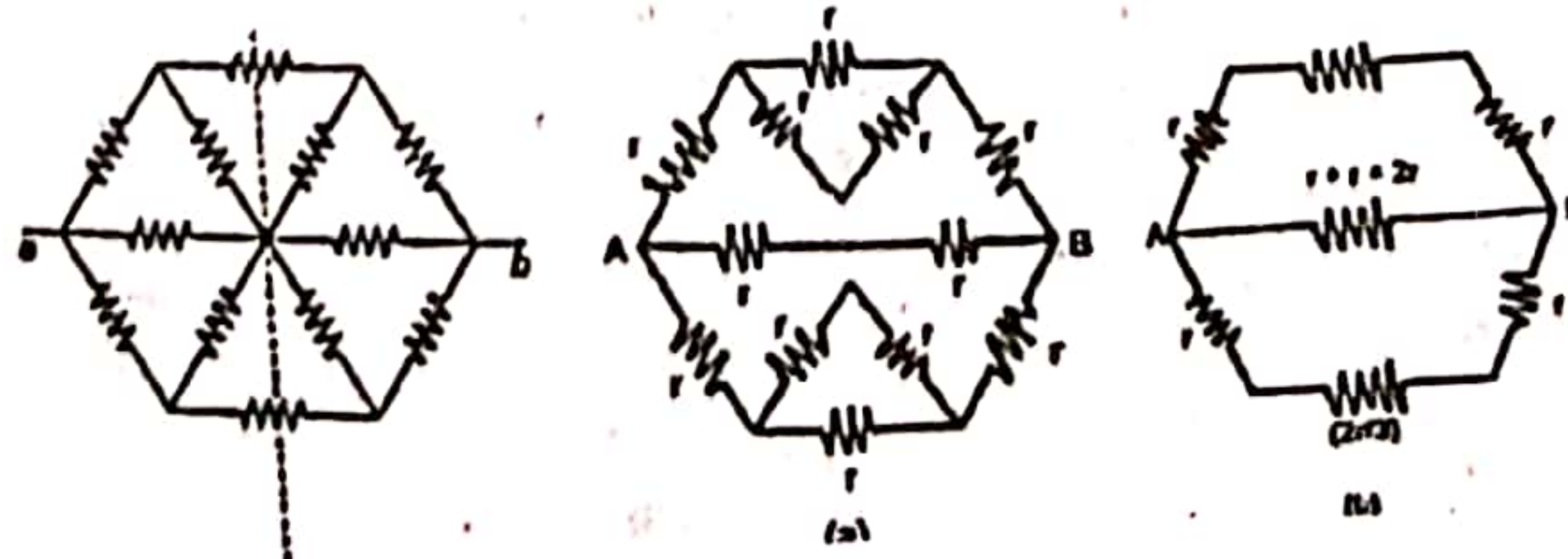
- (1) $\frac{V}{8R}$
- (2) $\frac{V}{4R}$
- (3) $\frac{V}{2R}$
- (4) $\frac{V}{R}$
- (5) $\frac{2V}{R}$



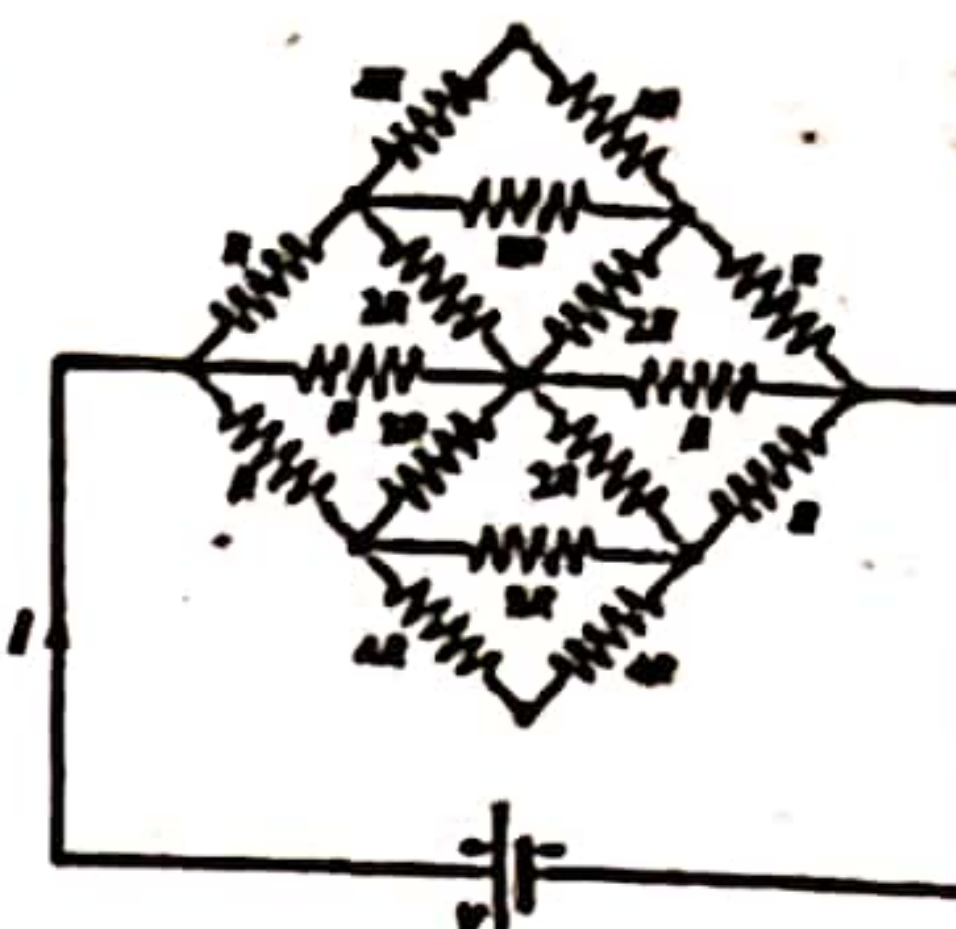
Ohm's Law combination of Resistance

08

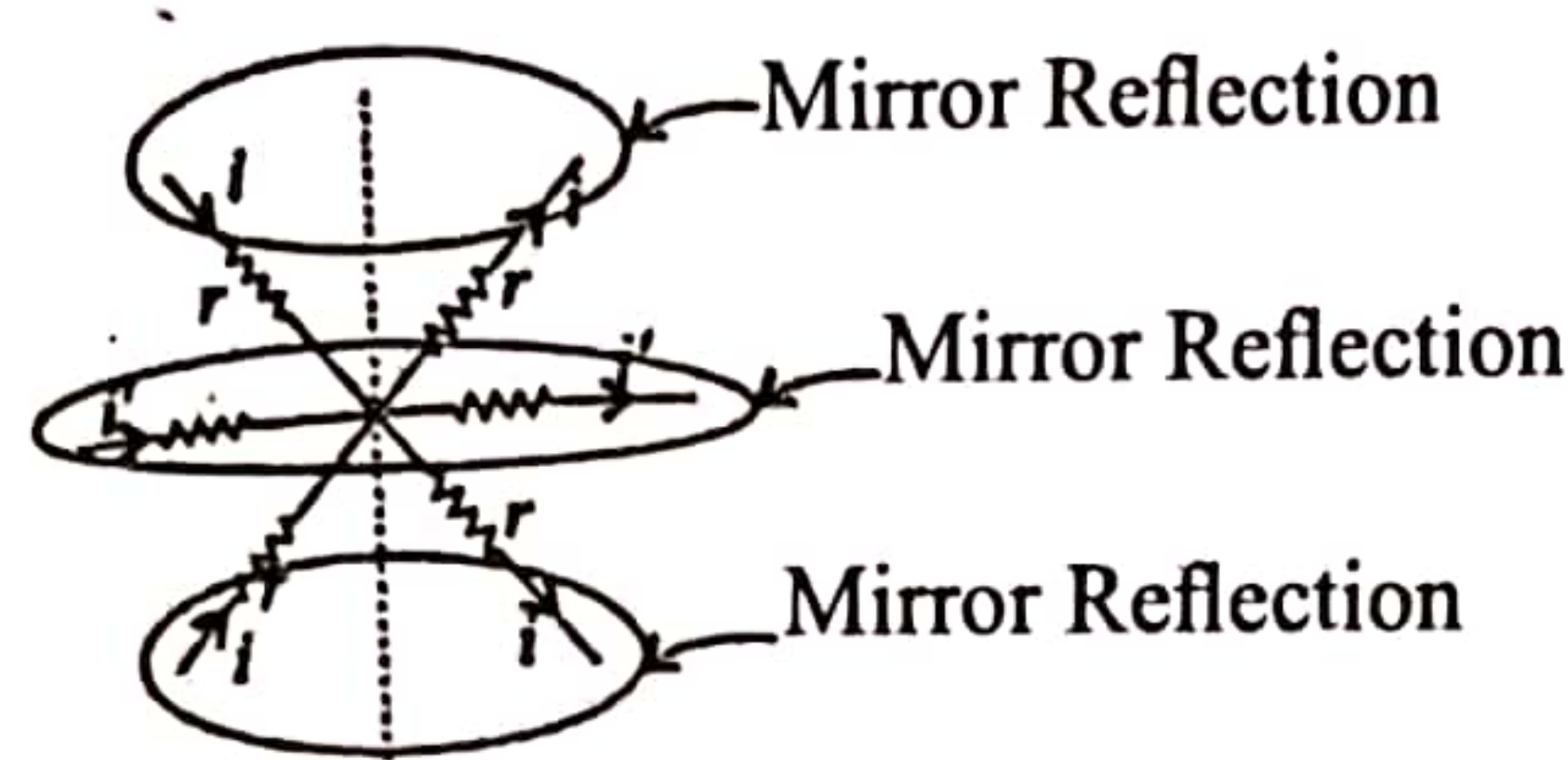
Consider the given resistor network according to the figure. The value of each resistor is r . We need to find the equivalent resistance between A and B. In such questions, always think of the symmetry. The resistor network is symmetrical around the vertical dashed lines as shown. It means that if we keep a mirror in front of the vertical dashed lines, then the image can be seen from both sides of each other. If you can find a symmetrical axis (line), then solving the problem is very easy. There are two rules which are associated with such symmetric axis.



- (1) The potential of any point on such a symmetric axis is equal. It indicates that the potential drop between any two points on the symmetric axis is zero. Therefore, there is no current flow across any resistor that has been connected along the symmetric axis. So, if there are such resistors, then without a problem you can remove them. This rule has no usage over the resistor network in the question. There are no resistors that are connected along the axis.



- (2) Equal current flows in the branches of resistors that are according to each mirror image around the symmetric axis. This rule can be applied to this resistor network. For example;



According to this reason, the resistor branches can be removed from the middle point. Figure (a) has shown this fact. There is no effect to the resistor network by loosening the middle joint. Now the rest of the part is easy. It is in series with r and r . For that $2r$, r is in parallel. The equivalent of $2r$ and r parallel arrangement is $2r/3$. You do the rest. The same logic can be applied to the given resistor network.

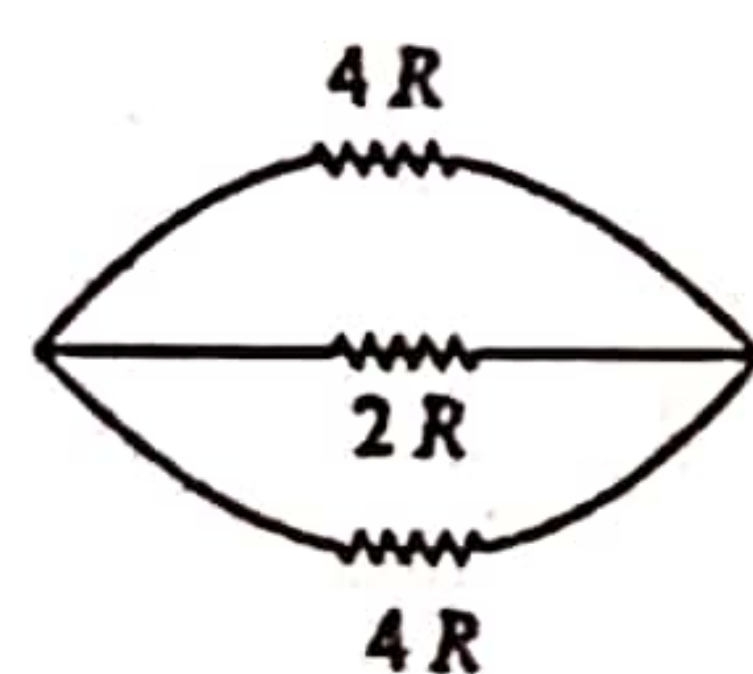
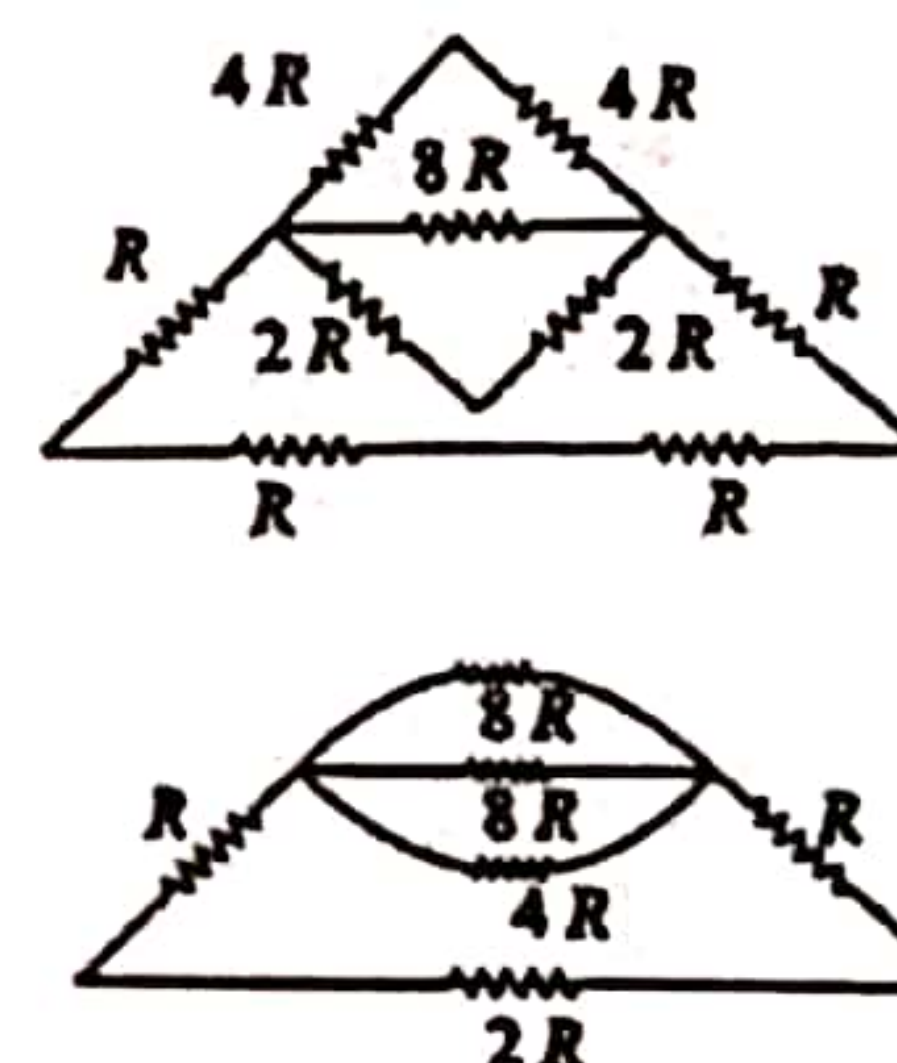
I have not drawn the bottom part. It is same as the top part.

$4R$ and $4R$ are in series. $2R$ and $2R$ are also in series. $8R$, $8R$ and $4R$ are in parallel.

$$1/R' = 1/8 + 1/8 + 1/4 = (1+1+2)/8 \rightarrow R' = 2R$$

You do not have to do the calculation like this way. You can do it from your memory. When 8 and 8 are in parallel, then the equivalent is 4 . When that 4 and the rest of 4 are in parallel, then the equivalent is 2 . The two R resistors that are on the sides of this $2R$ are in series. That means the equivalent is $4R$. There is another $4R$ drawn from the bottom.

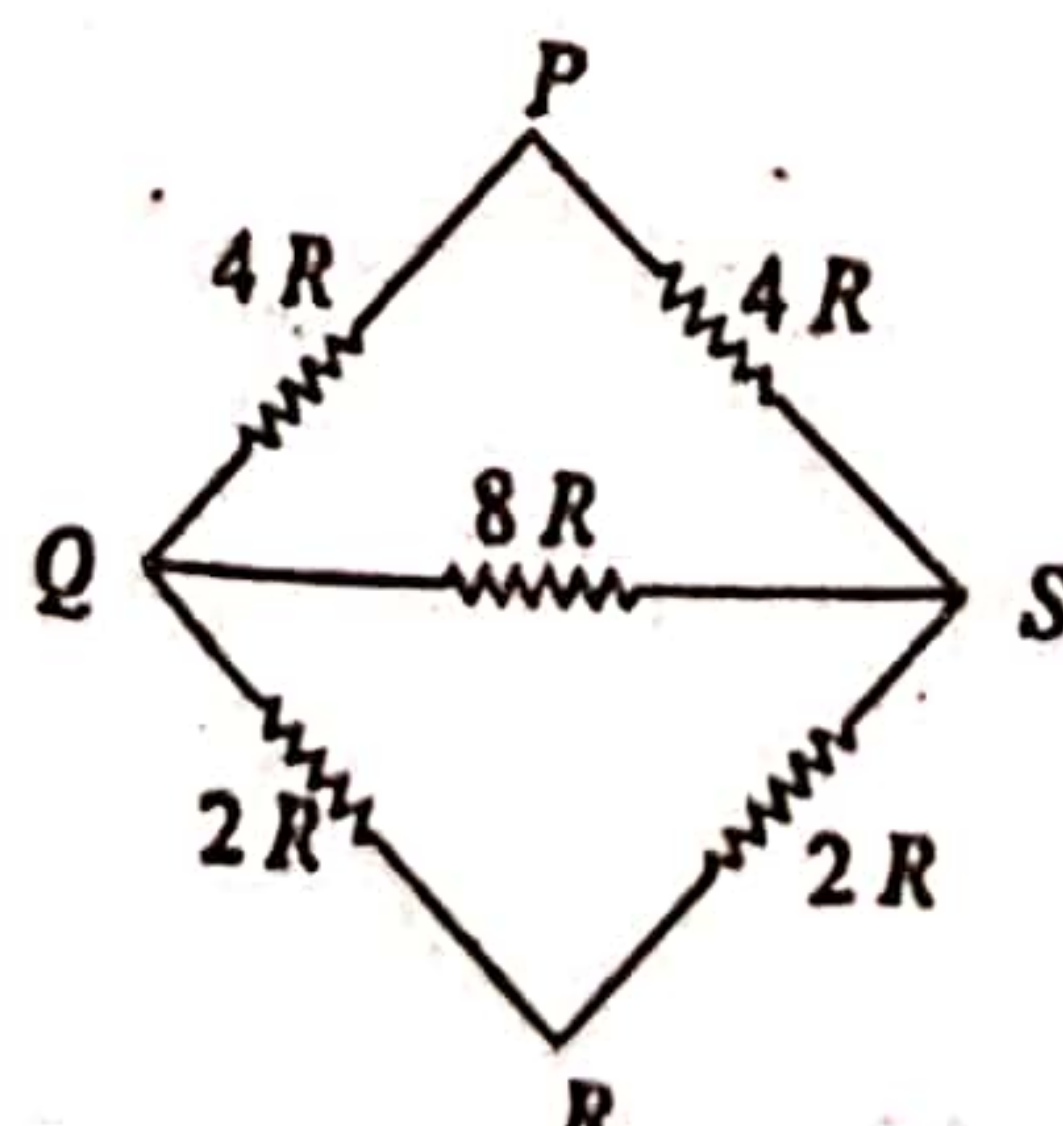
Now the final equivalent $1/R'' = 1/4 + 1/4 + 1/2 = 1$. This can be done quickly from your head. When 4 and 4 are in parallel, then the equivalent is 2 . Finally, when 2 and 2 are in parallel, then the equivalent is 1 . $R'' = R$



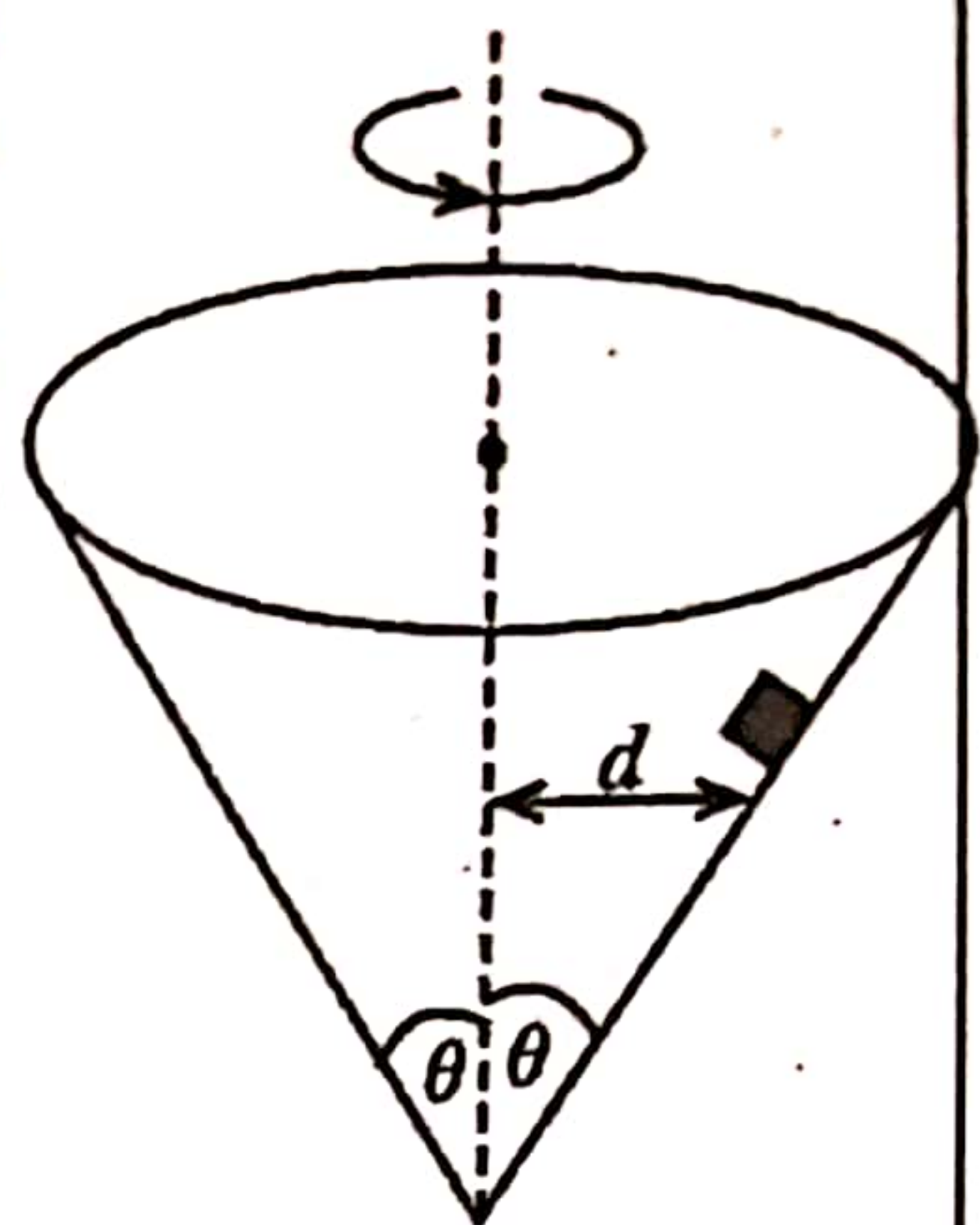
When two quarters and added to a half, the total is one. Therefore, the current that flows $= V/R$

Note: Do not remove $8R$ here. It is wrong. We know that $V_P = V_R$. But V_Q is not equal to V_S . There is a potential difference between points Q and S . If $V_Q = V_S$, then there will not be any current flow in any branch. If there is a connected resistor across P and R , then you can remove it. The potential difference gets zero only across the symmetric axis.

Keep the two rules in mind. From that you can solve many problems like these.



50. A small object is placed inside a right circular cone with axis vertical and vertex down as shown in the figure. The coefficient of static friction between the inner surface of the cone and the object is μ . What is the maximum angular velocity of rotation of the cone about its axis for the object to be on the inner surface of the cone without slipping at a distance d away from the axis?



(1) $\sqrt{\frac{g(\cos \theta - \mu \sin \theta)}{d(\sin \theta + \mu \cos \theta)}}$

(2) $\sqrt{\frac{g(\sin \theta - \mu \cos \theta)}{d(\cos \theta + \mu \sin \theta)}}$

(3) $\sqrt{\frac{g(\cos \theta + \mu \sin \theta)}{d(\sin \theta - \mu \cos \theta)}}$

(4) $\sqrt{\frac{g(\sin \theta + \mu \cos \theta)}{d(\cos \theta - \mu \sin \theta)}}$

(5) $\sqrt{\frac{g}{d \tan \theta}}$

Circular Motion

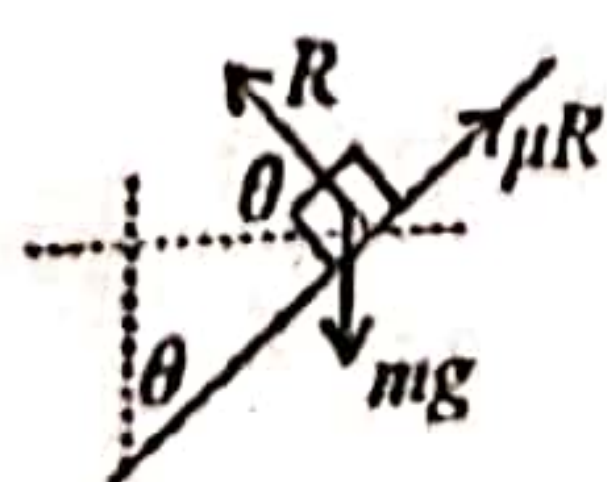
02

A small block with mass m is placed inside an inverted cone that is rotating about a vertical axis such that the time for one revolution of the cone is T . The walls of the cone make an angle β with the horizontal. The coefficient of static friction between the block and the cone is μ_s . If the block is to remain at a constant height h above the apex of the cone, what are (a) the maximum value of T and (b) the minimum value of T ? (That is find expressions for T_{\max} and T_{\min} in terms of β and h .)

This question can be seen in many books. There was such a question even in Olympiad question paper of 2016. First, you need to draw the forces acting on the object. It is easy to draw the perpendicular reaction (R) and the weight of the object (mg). To which side we should draw the frictional force? Is it upwards or downwards of the cone surface? Most of the time, the frictional force tends to draw upwards along the surface. That is because we think that the object may try to slide downwards. But we can consider two instances here. If the cone is rotated slowly, then the object can be kept at rest relative to the cone by stopping the motion of the object which will come down. However, if the cone remains without rotating, then the object tends to slide downwards.

Whether the cone is rotated or not, the object should be positioned without sliding on the surface. If it slides even before the rotation, then we cannot ask the question. Do you understand that to satisfy that condition μ should be $\mu \leq 1/\tan \theta$? Think about it for a while. The angle θ is measured from the vertical direction not from the horizontal direction.

But there is another instance. If the cone is rotated in a fast speed, then the object tends to go upwards along the surface. On that occasion, the frictional force is acted downwards along the surface. Therefore, the directions of the forces that act on the object can be represented like this way for two instances.



minimum angular speed



Maximum angular speed

The question asks about the maximum angular speed. Therefore, we need to pay attention on the second figure. Actually, the object is not at rest relative to the ground. It moves at a circle with a radius of d . If so, then there should be a net force that is directed towards the centre of the circle. Applying $\leftarrow F = ma$, $R \cos \theta + \mu R \sin \theta = md\omega^2 \dots (1)$ (as ω is asked)

I have written the centripetal acceleration as $d\omega^2$ ($r\omega^2$) not as v^2/d (v^2/r). The resultant of the vertical forces acting on the object should be zero. That is because there is no acceleration towards that direction.

$$\uparrow R \sin \theta - \mu R \cos \theta = mg \dots (2)$$

$$(1)/(2) \frac{\cos \theta + \mu \sin \theta}{\sin \theta - \mu \cos \theta} = \frac{d\omega^2}{g} \rightarrow \omega = \sqrt{\frac{g}{d} \cdot \frac{\cos \theta + \mu \sin \theta}{\sin \theta - \mu \cos \theta}}$$

When equations are written for the minimum angular speed instance, then you will get these.

$$\leftarrow R \cos \theta - \mu R \sin \theta = md\omega_1^2$$

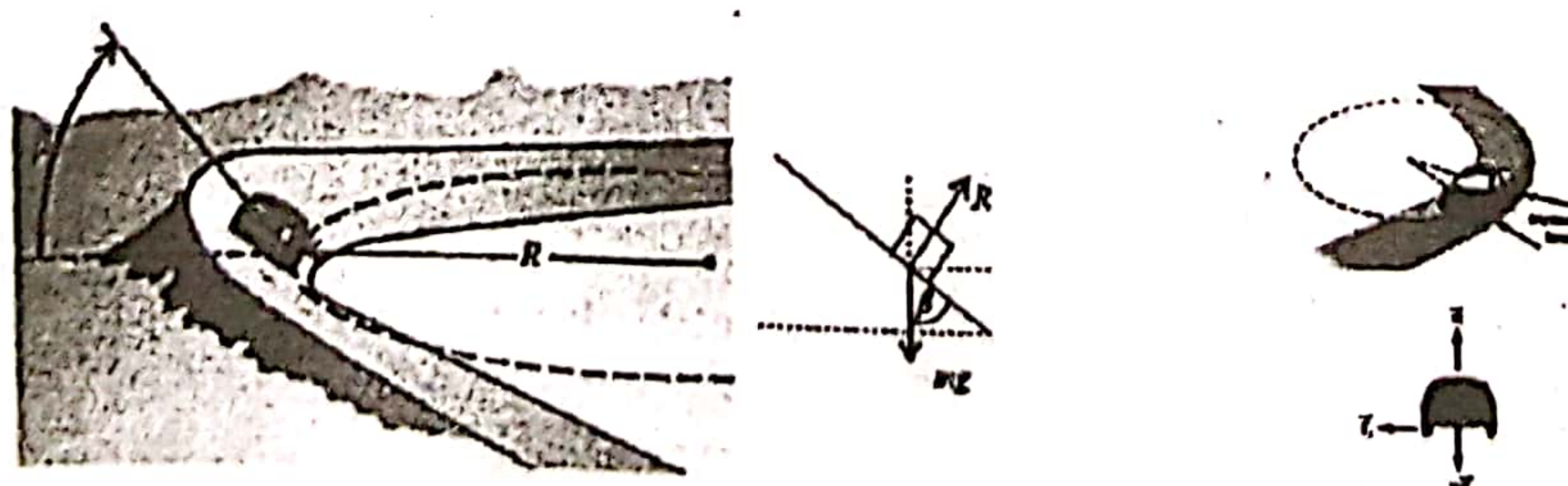
$$\uparrow R \sin \theta + \mu R \cos \theta = mg$$

$$\text{From this you will get, } \omega = \sqrt{\frac{g}{d} \cdot \frac{\cos \theta - \mu \sin \theta}{\sin \theta + \mu \cos \theta}}$$

Hope many children have gone for this answer. When the speed is increased, why does the object tend to go upwards along the surface? When the speed is high, the force which is directed towards the centre also should be increased. To satisfy that, the component of the frictional force should be directed towards the centre. $R \cos \theta$ is however acting towards the centre. If you need to add another force, then it can be supplied only from the frictional force. However, mg is a vertical force. Initially, the resultant force is $\leftarrow R \cos \theta - \mu R \sin \theta$. If you need to increase the force, then there is no other alternative for nature/ Physics than changing $-\mu R \sin \theta$ into $+\mu R \sin \theta$.

When a bus takes a bend, if someone argues that we tend to press the other passenger to find the extra force not for the desire, then I do not see something wrong in that. If $\mu=0$, then $\omega = \sqrt{\frac{g}{d} \cdot \frac{\cos \theta}{\sin \theta}}$

According to the figure, when a vehicle is taking a bend in an inclined path, this is the expression for the speed of the vehicle. Here, the component of R obtained from the inclined plane without friction provides the needed centripetal force. This is the same relation as above.



$$\rightarrow R \sin \beta = \frac{mv^2}{r} \text{ and } \uparrow R \cos \beta = mg. \text{ Then } \tan \beta = \frac{v^2}{rg}$$