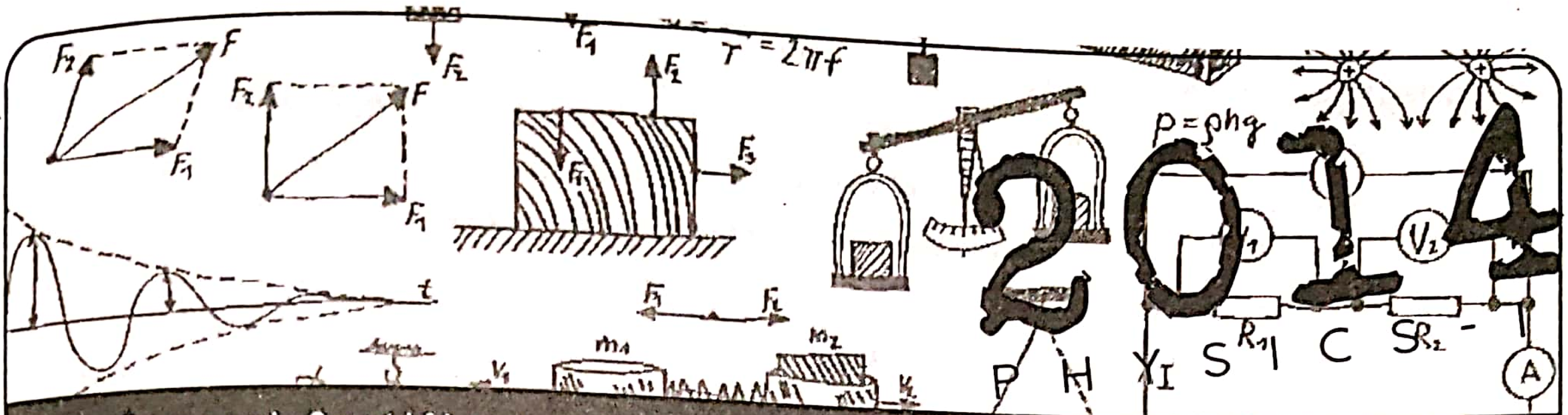
SCIENCE

@ALSCIENCESTUDENT
SDISCUSSIONGROUP



1. As far as the units are concerned, which of the following quantities differs from the rest?

(1) Rotational kinetic energy (2) Mechanical potential energy (3) Internal energy
(4) Work (5) Power

Unit and Dimensions 01

The first three answers are energies. The 4th answer is work. The unit of work is also equal to the unit of energy. There should be energy to do a certain work. Power is the rate of work done. Its unit is W (Js^{-1}).

2. Which of the following quantities is/are dimensionless?

(A) Relative Velocity

(B) Relative density

(C) Relative humidity

(1) A only. (2) A and B only. (3) B and C only
(4) A and C only (5) All A, B and C.

Unit and Dimensions 01

You can just flow till the 8th question. There is dimension for relative velocity. But there are no dimensions for relative density and relative humidity. Relative density means the determination of a density of a certain material or liquid in an amount/ fraction compared to the density of water. For example, the relative density of mercury is 13.6. What is meant here is that the density of mercury is 13.6 times greater than the density of water.

Water is a known liquid by us. It is essential for living. Therefore, considering other densities relative to the density of water is fair. In our life also, an unknown thing can be expressed using a known thing. For example, we say that, 'look that person is like Sharukh Khan or else like Aishwarya Rai'. In our younger days we said specific gravity for relative density. Now this term is not in use. We know from the interpretation of relative humidity that it does not have units or dimensions.

3. Which of the following propagates in the form of longitudinal waves?

(1) Laser light (2) X-rays (3) Ultrasonic waves
(4) Microwaves (5) Radio waves

Wave Properties 03

You can get the answer as soon as you see. Except the ultrasound waves, all the others are electromagnetic waves. The electromagnetic waves are transverse waves. Even there are frequencies that exceed our auditory range, the ultrasound waves are a form of waves.

4. When a guitar is played, it will produce

(1) longitudinal progressive waves on the strings and longitudinal progressive waves in air.

(2) transverse progressive waves on the strings and longitudinal progressive waves in air.

(3) longitudinal standing waves on the strings and transverse progressive waves in air.

(4) transverse standing waves on the strings and longitudinal progressive waves in air.

(5) transverse standing waves on the' strings and transverse standing waves in air.

Transverse Waves

You can realize that the answer is (4) once you read the question. When a guitar wire is vibrated, it is vibrating under two limits. One displacement node is created on the balance whereas the other is created on the place where it is tightened with the finger most of the time. Therefore, transverse standing waves are being created on the wire. Longitudinal progressive waves are being created in the air from the vibration of the wire. There are no boundaries to the wave propagation in the air. They vibrate your ear drum as well as my ear drum.

5. Which of the following statements is not true with regard to a compound microscope?

(1) It has two convex lenses.

(2) Image. of the object formed by the objective is real.

(3) Separation of the lenses is much greater than the focal length of the objective or the eyepiece.

(4) Final image fanned by the microscope is a virtual image.

(5) The object to be examined should be placed within the focal length of the objective

Opticle Instrument

You can understand that the false statement is (5). When the object is kept at the focal length of the objective, then the image created on the objective is unreal. If so, the work is hard. Even the final image can be unreal but the first image from the objective should be a real one. The object should be placed away from the focal length of the objective. Then the image from the objective will be created away from $2f_o$ (f_o is the focal length of the objective). From this you can see that 3rd sentence is correct. The real image from the objective should be placed in the focal length of the eyepiece.

6. Two cells, each having e.m. E and internal resistance r; connected as shown in figure are equivalent to a single cell with

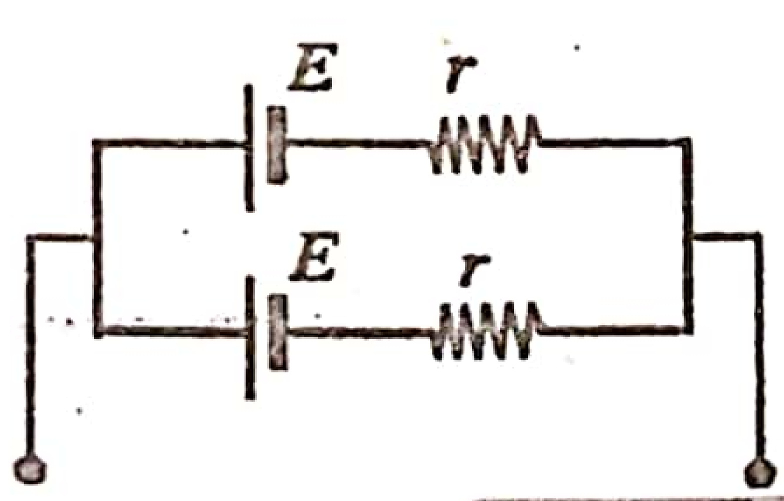
(1) e.m.f. E and internal resistance r

(2) e.m.f. 2E and internal resistance 2r

(3) e.m.f. 2E and internal resistance r

(4) e .m .f. E and internal resistance $\frac{r}{2}$

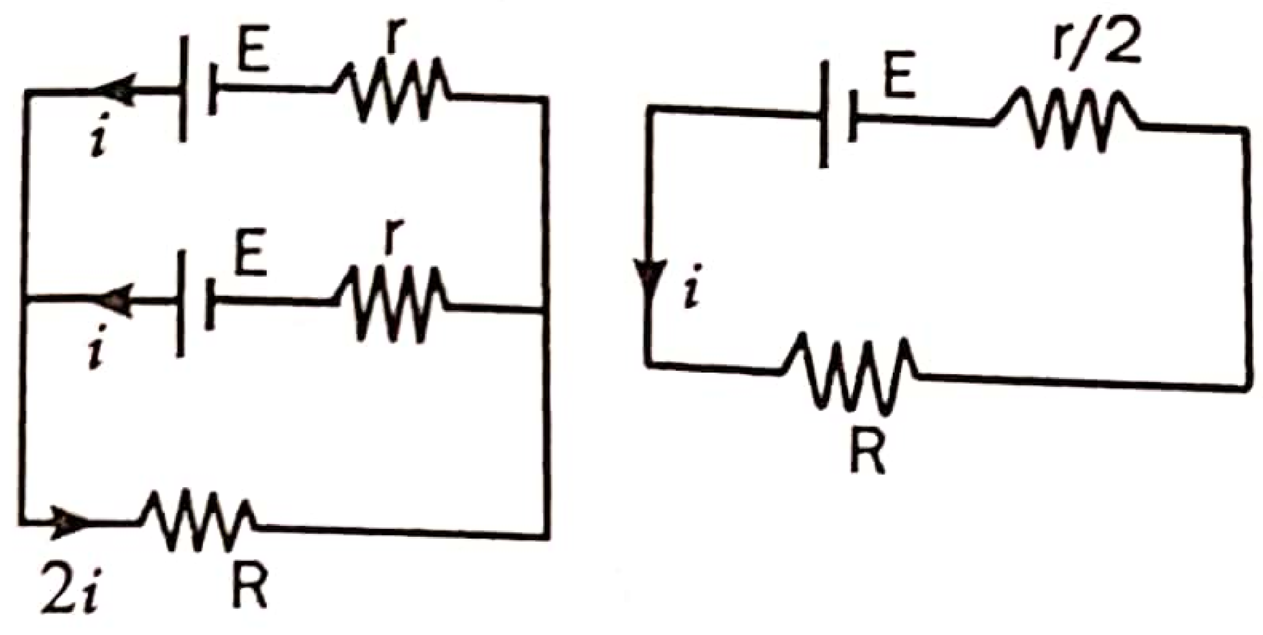
(5) e.m.f. E and internal resistance 2r.



Korchoff's Law- Combinations of Cells

It is very simple and a known question. When number of e. m. f of E are in parallel, the equivalent e. m. f is also E. When two equal resistors are in parallel, their equivalent resistance is half of it. The advantage of cells being parallel is that it allows current flow across an external resistor for a longer period. The internal resistance of the cell is indicated as r. It is not a resistance that is being connected from outside. The internal resistance cannot be drawn inside the cell. Normally it is not drawn like that. Therefore, it is wrong to think that there is another r for the cell apart from the drawn r. From the following calculation, you can understand that the equivalent circuit is correct.

Two metals of A and B were subjected to photoelectric effect where metal A has the highest work function than that of metal B. Which graph is correct?



E = ir +2iR; i= E/(r +2R)

The current across the external R = 2i = 2E/ (r +2R)

E = i.r/2 +iR

The flowing current i= 2E/ (r +2R)

Relative to an outsider, both do the same thing.

7. Two charged conducting spheres of radii $R_1 = r$ and $R_2 = 2r$ are connected by a thin conducting wire. After being connected, if the respective charges on the two spheres are Q_1 and Q_2 and, the corresponding surface charge densities on the two spheres are a_1 and a_2 respectively, then

(1) $\frac{Q_1}{Q_2} = \frac{\sigma_1}{\sigma_2} = \frac{1}{2}$ (2) $\frac{Q_1}{Q_2} = \frac{\sigma_1}{\sigma_2} = 2$ (3) $\frac{Q_1}{Q_2} = \frac{1}{2}, \frac{\sigma_1}{\sigma_2} = 2$

(4) $Q_1 = Q_2, \sigma_1 = \sigma_2$ (5) $\frac{Q_1}{Q_2} = 2, \frac{\sigma_1}{\sigma_2} = \frac{1}{2}$

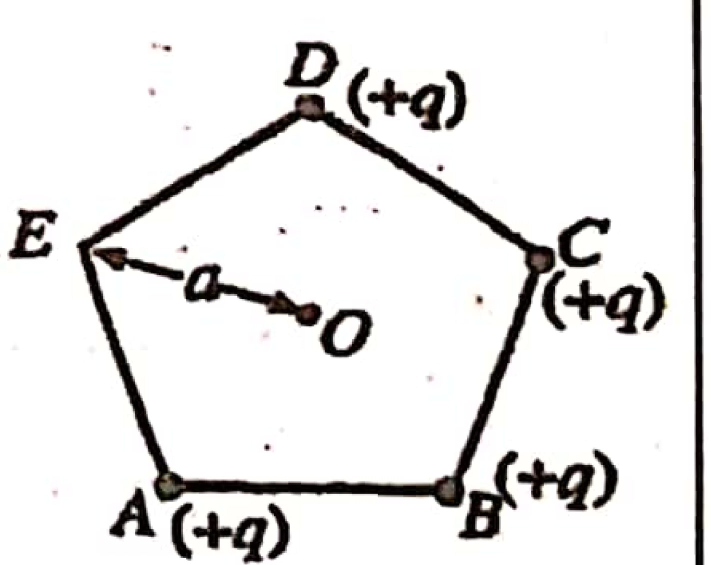
Electrostatic Potential 06

When it is being connected from the wire, the potentials of the spheres are equal. That means $Q_1/r = Q_2/r$. There is no need to write $1/4\pi\varepsilon o$. $Q_1/Q_2 = ½$. Only (1) and (3) are being left. Surface charge density means the amount of charge in a unit area of a surface. $Q_1 \propto \sigma_1 r^2; Q_2 \propto \sigma_2 4r^2$

Therefore, $\frac{\sigma_1}{4\sigma_2} = Q_1/Q_2; \frac{\sigma_1}{\sigma_2} = \frac{4}{2} = 2$

I will first do my rough work for this question. Even without rough work you can get the answer by logic. To have the same potential, the charge of the sphere with two times radius should be double compared to the charge in the sphere with a radius. That means $Q_2 = 2Q_1$. Next as $\sigma_1 r^2$ and $Q_2 \propto \sigma_2 4r^2, 4\sigma_2 = 2\sigma_1$.

8. Four particles each having a charge of +q are placed on four vertices of a regular pentagon as shown in figure. The distance from the centre O of the pentagon to a vertex Is a. The electric field Intensity at the centre of the pentagon is

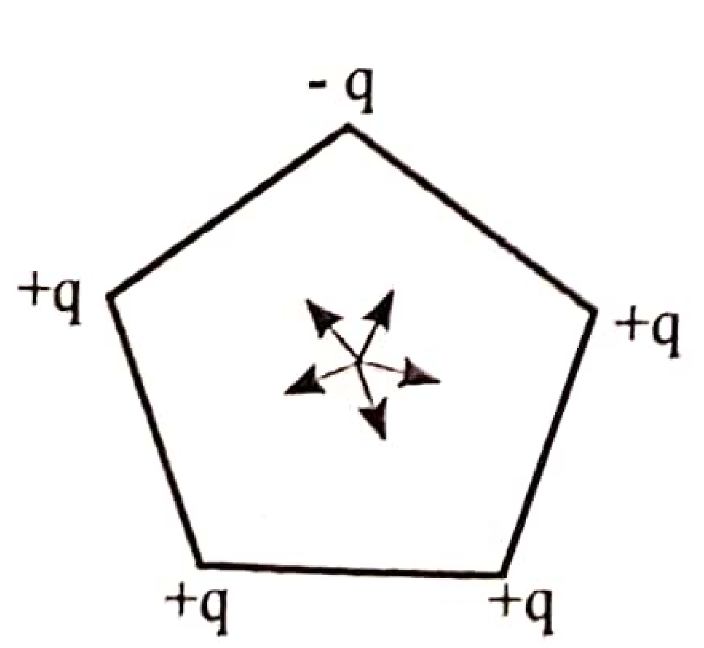


(1) $\frac{q}{4\pi\varepsilon_0 a^2}$ in the *OE* direction. (2) $\frac{q}{4\pi\varepsilon_0 a^2}$ in the *EO* direction.

(3) $\frac{q}{\pi\varepsilon_0 a^2}$ In the *OE* direction. (4) $\frac{q}{\pi\varepsilon_0 a^2}$ In the *EO* direction.

(5) zero.

Electric Field Intensity and Coulomb's Law 06



If you do not put logic, then you will be at a loss. If you find field intensities from the four charges and try to find the resultant, then as there is nothing to do, you will blame the people who made the paper. Here is the logic. If equal charges were there in 5 points, then the electric field intensity of the centre O should be zero. Due to the symmetry, this fact can be obtained without a calculation.

But there is no charge in the point of E. Therefore, the resultant electric field intensity from the rest of four charges should be equal and opposite to the electric field intensity at the centre if there was a charge of +q in the point of E. If this conclusion does not work, then the electric field intensity at the centre O cannot be zero when there are five +q charges at the corresponding points. If there is a +q charge at point E, then the electric field strength at O from it is towards $\frac{1}{4\pi\varepsilon_0}\frac{q}{a^2}$ EO direction. Therefore, the electric field strength at O from other +q charges should be $\frac{1}{4\pi\varepsilon_0}\frac{q}{a^2}$ towards OE direction.

One can argue that it is better if this question was given at the end of the paper instead of giving at the start. But the examiners have given this question to emphasize that it is not a complex question. I feel that majority of students have marked the direction of each filed intensities and finally wasted a lot of time by finding their resultant (which is very difficult).

If you find that this cannot be solved by the above way, then you need to find another way. You must understand that there cannot be such a time-consuming question for MCQ. Then you should have looked at the answers. You know that the electric field intensity cannot be zero at O (from the four charges). There is the direction OE or EO in all the four answers. There is no other direction. From this you need to catch the logic. You need to be intelligent to catch the logic. Examiners are not cruel. Now (3) and (4) can be removed. Why? Because there is no 4 in $4\pi\varepsilon_0$ As there are four charges, you can think that 4 is cut off from $\frac{1}{4\pi\varepsilon_0}$ when multiplying by 4. But as the electric field intensities from four charges are acting on different directions, you should realize that you cannot just multiply by 4. Then what is left is (1) and (2). If you pick with the blind sight, then you should choose either (1) or (2). There were some students who realized this logic and picked EO by mistake. That is unfortunate. The given choices are enough to emphasize that the problem is tangled with EO or OE directions. If the logic has come to your brain quickly, then you are lucky. How much time did students waste for this question? If you did not get the above logic, then you cannot get the answer. But do not waste the time. Either abandon or pick a blind choice and go to the other questions.

9. **A thin ring of mass M and radius R is rotating in a horizontal plane about an axis passing through its centre perpendicular to its plane with a constant angular velocity ω. Now if two small masses, each of mass, are attached gently to the opposite ends of a diameter of the ring, the new angular velocity of the system is**

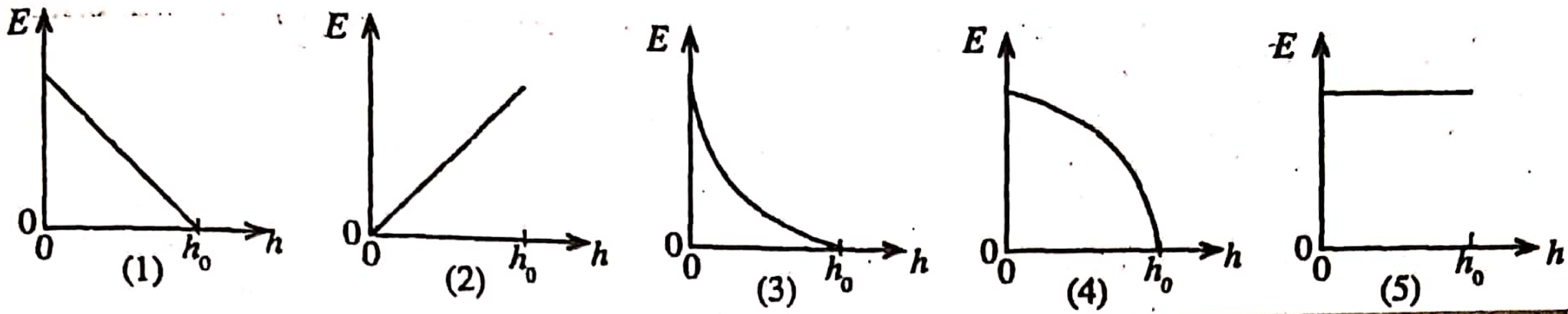
(1) $\frac{\omega M}{M+2m}$ (2) $\frac{\omega(M+2m)}{M}$ (3) $\frac{\omega M}{M+m}$ (4) $\frac{\omega(M-2m)}{M+2m}$

Rotational Motion

It is a simple question. It is not essential to draw a figure for this question. Once you read this, you need to understand that conservation of angular momentum should be applied. As the masses were connected carefully and the new angular velocity of the system is being asked, you need to realize that you should apply conservation of angular momentum. It has been mentioned that the masses of m were carefully inserted to emphasize the fact that there is no external torque or there is no damage to the system when they are being connected.

You should know the moment of inertia around a perpendicular axis across the centre of a thin ring. As the ring is thin, that value is MR^2. As the masses are also small, the moment of inertia of a mass around a perpendicular axis across the centre is mR^2. You need to know mR^2. You do not have to know the equations for moment of inertia for other structures like disks, rods etc. There is no need to know the equations of moment of inertia. Apply conservation of angular momentum directly. $MR^2\omega = MR^2\omega' + 2mR^2\omega'$; $\omega' = M\omega/(M+2m)$. If the figure was created in your mind, then you can get the answer even without rough work. Avoid drawing figures as it is a waste of time.

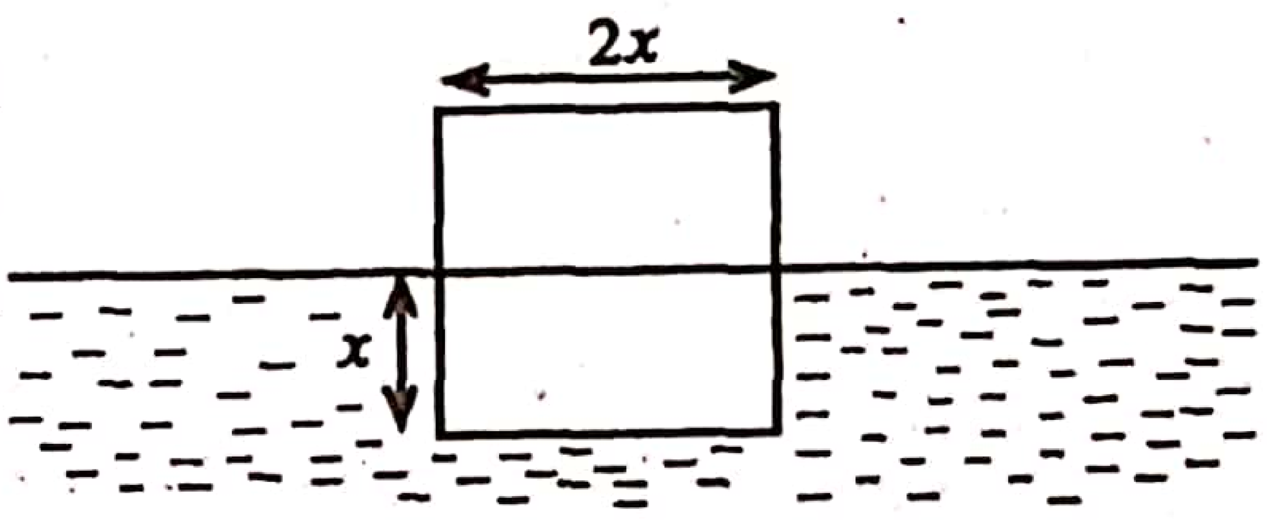
10. A particle of mass m is dropped freely from a location at a height ho from the ground. The variation of the kinetic energy (E) of the particle with height h as measured from ground is best represented by



Work Power and Energy 02

It is simple. I can remember it was there in a past paper. The potential energy is zero at h_0 height. Why? It was dropped freely. From this (2) and (5) choices are out. When h is reduced from h_0, the kinetic energy increases. When it comes downwards, the kinetic energy increases. As the total of kinetic energy and the potential energy is constant, the variation of the kinetic energy with h cannot be a curve. That means the answer is (1). If you write an equation, in a height of h up from the ground, then E + mgh = constant. That means E = - mgh + constant. So, the variation of E with h should be a straight line with a negative gradient. The variation of h with velocity v is a curve. But in the y axis it is drawn the kinetic energy not the velocity.

11. A solid cube of plastic of mass M and side length 2x floats in water with half the side length submerged as shown in figure. If this cube is now converted into a hollow cube of mass M with external side length 8x, the depth to which it submerges in water will be



(1) $\frac{x}{2}$ (2) $\frac{x}{4}$ (3) $\frac{x}{8}$

(4) $\frac{x}{16}$ (5) $\frac{x}{32}$

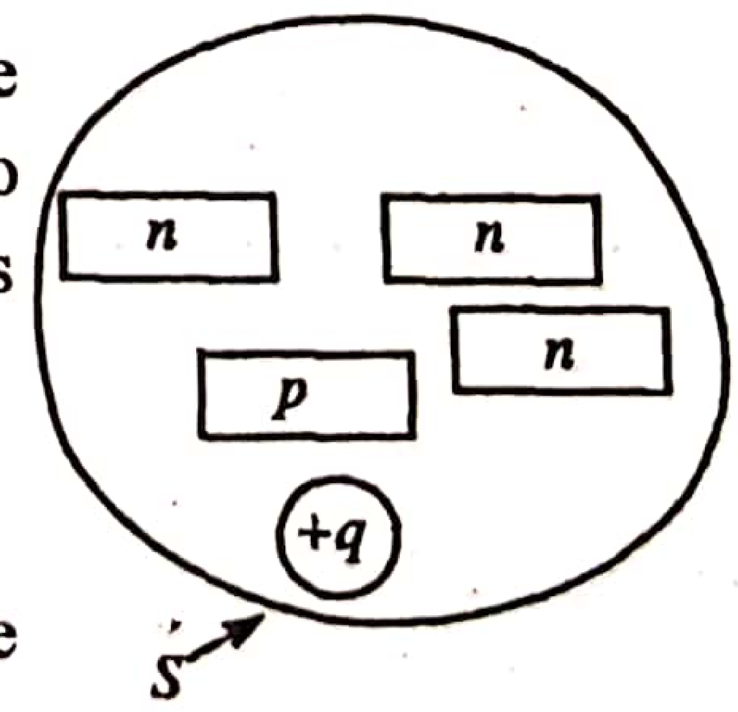
Hydrostatics 02

You do not have to waste much time on this question. There is no change in the mass of the cube. Therefore, in both instances, the acting upthrust should be same. Even they are being sunk in same water. Therefore, the sunk volume of both occasions should be same. Now you can get the answer by one equation. If the sinking height of the second instance is y, then

2x. 2x. x = 8x. 8x. y; y= x/16

You know that each length of a cube is equal in Ordinary Levels. Therefore, the area of a side of the cube initially is 2x. 2x = 4 x^2. Later it has been changed into 8x. 8x = 64 x^2. Initially, the volume of the cube is 8 x^3. The volume afterwards is 8x. 8x. 8x =512 x^3. So, to keep the weight unchanged, the middle of the cube should be eaten.

12. A Gaussian surface S encloses a metal sphere carrying a charge of +q, three n-type semiconductor pieces each having a number of free electrons corresponding to charge of - q, and one p-type semiconductor piece having a number of holes corresponding to charge of +q as shown in figure.

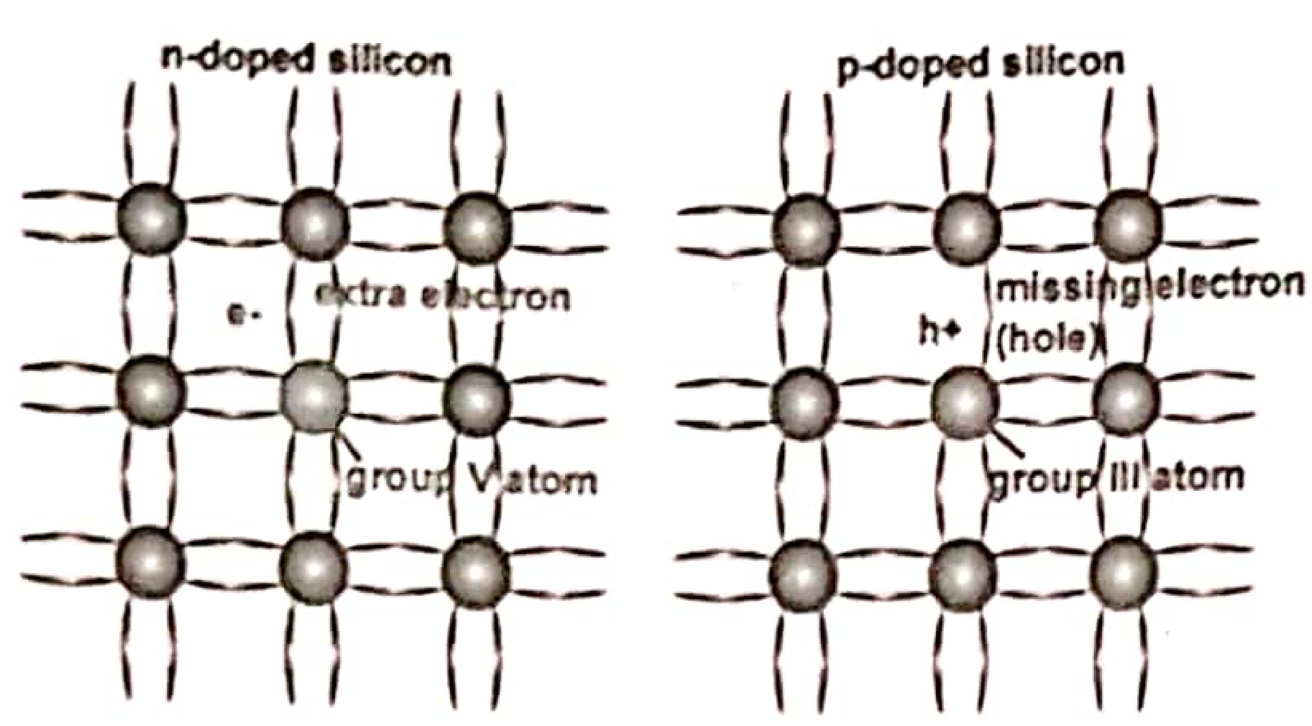


Total electric flux through the surface can be made zero by
(A) removing one n-type semiconductor piece.
(B) adding one more p-type semiconductor piece with the same hole concentration.
(C) bringing a metal sphere carrying a charge of - q from outside into the enclosed volume.

Of the above three methods
(1) only A is true. (2) only C is true.
(3) only A and B are true. (4) only B and C are true
(5) All A, B and C are true.

Gauss Theorem 06

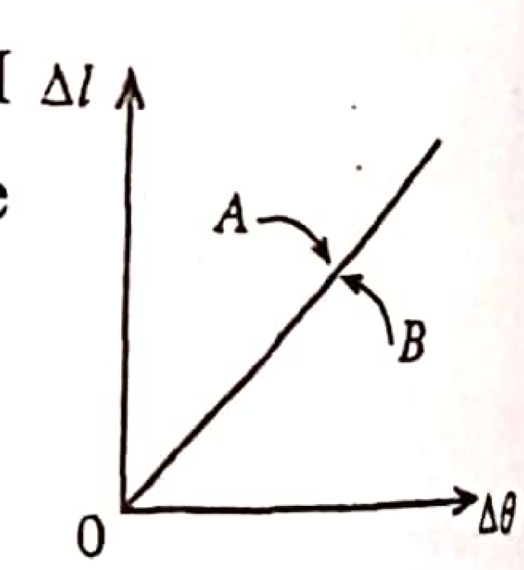
I suspect that the answer for this question was dropped from many children. There is no net charge in n type or p type semiconductors. You know that n type semiconductor is made when Silicon lattice is being doped by a pentavalent As, Sb or P. Think that in the lattice of Silicon there is a substituted As atom instead of Si atom. Look at the figure.



In the last valent shell of As atom, there are five valence electrons. Out of them, four are bonded with the nearby four Si atoms. The left one electron of As atom is there as a free electron without being used for bonding. In the question what is mentioned as each free electron corresponds to a charge of -q means that these electrons where they remain as free electrons released from As atoms. But there is no net charge in the semiconductor even there are certain number of free electrons.

When the electron is freed from As atom, it gets positively charged (As^+). Look at the figure. Therefore, positive and negative charges (the negative charge in the free electron) are being cancelled off with each other. As atom has given an electron that is within itself. It is not an electron from the outside. There is no issue in the sentence fragment of number of free electrons corresponding to -q charge in n type semiconductor. Actually, in a n type semiconductor there are free electrons corresponding to -q charge. But there is no net charge on the semiconductor. Free and net means two different things not one. Even if they behave freely, as there are people waiting to be connected (positively charged As atoms) the final net is zero. We are also like this. Some argue that it is better if the above sentence was not there. Even one can build an argument that, it is just enough to give as n type and p type semiconductor pieces. It is true but the sentences are correct. There is no issue in the question. Simply, an electric flux will not be created from n type or p type semiconductors because they do not have a net charge inside. Therefore, forget about all n and p semiconductors. Then what is left will be the sphere with +q charge. So, statement (C) is only correct. You need to bring -q from outside to eat +q. If you put -q and +q to n and p respectively, then (3) will be correct. I do not know how many got (3) as the correct answer.

13. When two metal rods A and B at room temperature are heated together and their M expansions Δl are plotted with the increase in temperature ΔB the two curves are found to coincide with each other as shown in figure. This could happen only if



(1) the two rods are made of same material.
(2) length of A is same as the length of B.
(3) linear expansivity of A is same as that of B.
(4) the product 'linear expansivity x original length' is same for both rods.
(5) the two rods are heated together.

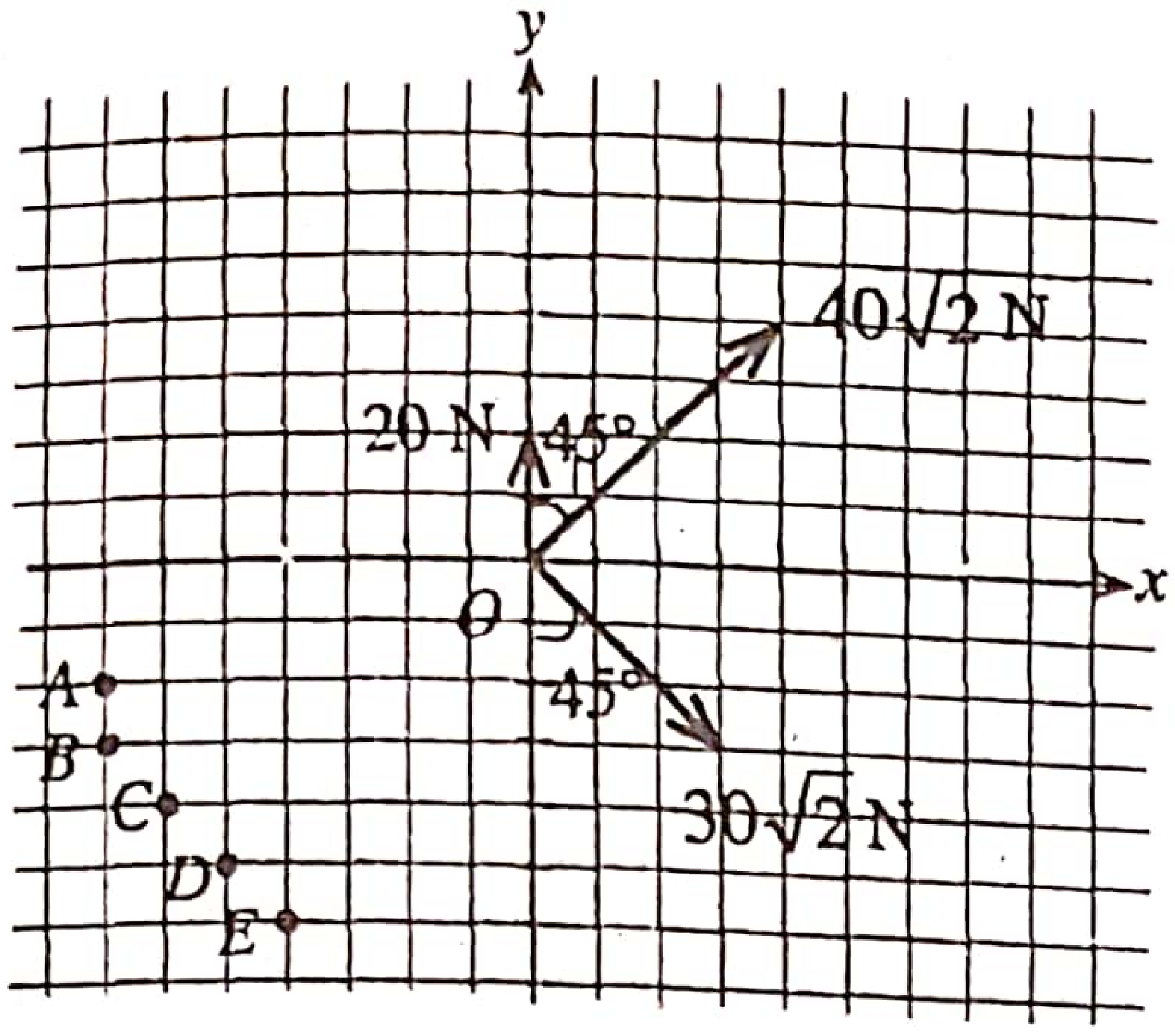
Expansion of solid

There is no game here. You know that $\Delta l = l_0 \alpha \Delta\theta$ by heart. If the same graph has to be there, then the multiple of l_0 and α must be equal. It is very simple.

14. If three coplanar forces of 20 N, $40\sqrt{2N}$ and $30\sqrt{2N}$ act on a particle situated at the origin O of a x-y coordinate system as shown in figure, the vector that represents the force necessary to keep the particle stationary is

(1) OA (2) OB (3) OC
(4) OD (5) OE

Equilibrium of Forces

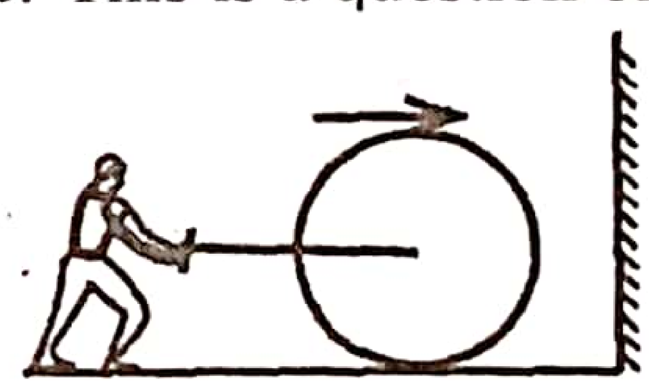


You can get the answer by drawing on the graph in the paper. There are couple of questions in the paper like this. For example, (41) and (42). First resolve $40\sqrt{2}$. As $\cos 45^0$ is $1/\sqrt{2}$, is , you can do everything from your memory. The vertical component of $40\sqrt{2}$ is 40 $(40\sqrt{2} \times \frac{1}{\sqrt{2}})$. Actually, you do not have to do like this. All have been drawn according to a scale. The vertical component of has 4 squares to the upward direction whereas it has 4 squares (to the x direction) to the horizontal direction. Likewise, $30\sqrt{2}$ has 3 squares to the bottom and 3 squares to the x direction. Now the resolving part is over. To the upward direction, there are 4 squares from $40\sqrt{2}$, 2 squares from 20 N and the total is 6 squares. To the downward direction, there are 3 squares from $30\sqrt{2}$. So, the resultant from the vertical components is 3 squares upwards (6-3). The total of horizontal component is 7 squares (4+3).

Therefore, the resultant of these forces is situated as 7 squares to the +x direction and 3 squares to +y direction. So, to keep the particle still, there should be an opposite and equal force. That means 7 squares to – x direction and 3 squares to – y direction. The point is B with those co-ordinates. The direction is OB. There is no need to draw the direction of the forces. It is enough to move your eyes from up and right and then move down and left.

Actually, the magnitude of the forces or the angles should not have been given for this question. It is enough if arrows are given. All have been drawn properly like a baby. What else is needed? This is a question of counting squares.

15. A heavy roller of mass 500 kg, moving on a horizontal. surface at a constant velocity of 1 m s^{-1} as shown in figure is stopped in 0.5 s on hitting a smooth vertical wall. The horizontal force exerted by the roller on the wall is



Newton's Laws and Momentum **02**

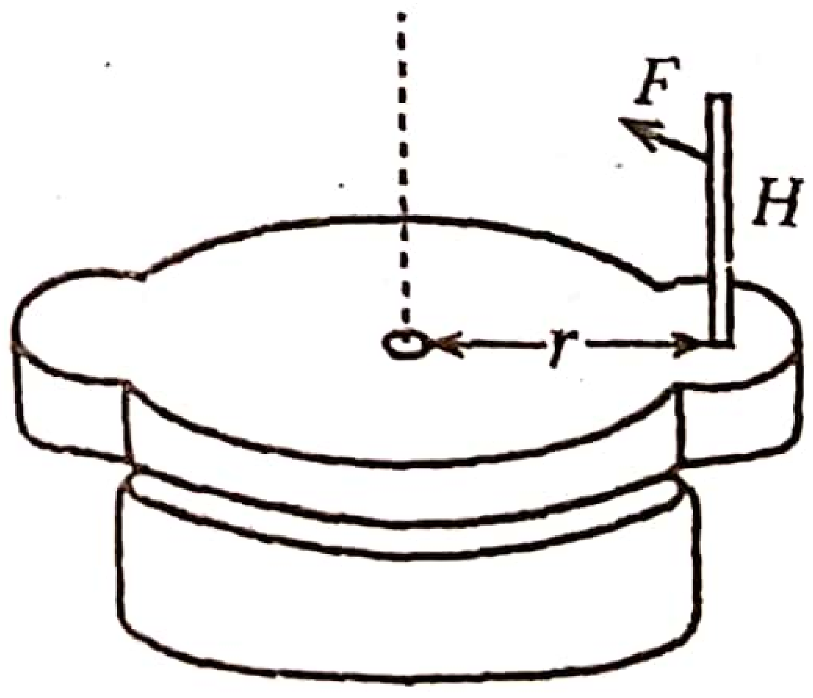
(1) 5 000 N (2) 3 000 N (3) 2 000 N

(4) 1 000 N (5) 500 N

It is simple as peanuts. You can do the calculation in seconds. The force is equal to the rate of change of momentum. In how many questions that this has been included?

(500 X 1)/0.5 = 1000 N

16. A traditional grain. grinder consists of two flat stones. The upper stone is rotated on top of the lower stationary stone by applying a horizontal force of magnitude F to the handle H which is fixed at a distance of r from the axis of rotation as shown in figure. If the force is always applied parallel to the direction of motion of the handle, and the period of rotation is T, the power being expended is



(1) $\frac{\pi r F}{T}$ (2) $\frac{2\pi r F}{T}$ (3) $\frac{rF}{T}$

(4) $\frac{F}{\pi r^2 T}$ (5) $\pi r^2 F T$

Rotatioal Motion **03**

This is also peanuts. When a torque of τ is rotated in θ radians, then the work done is $\tau\theta$. Here the moment of the applied F force is Fr. The angle that rotates in a rotation is 2π radians. Power is the energy per unit time. The time taken for one revolution is T. So, the power is $2\pi rF/T$. If needed, you can get the answer from the work done by the force. In one revolution, the work done by force F is $F2\pi r$. That means F is multiplied by the circumference. When it is divided by T, you will get the power. Actually, this is the minimum power that must be used. The time taken to get the answer is less than time taken to read the question. You can get the answer without doing rough work.

If you know the moment of inertia of the upper stone, expressions for things like the anguälr acceleration of the stone, the angular velocity of the stone after one revolution can be obtained. Try out and see.

17. A radioactive material has a half life of 60 minutes. The percentage of the fraction of material that has decayed during a period of 3 hours is

(1) 8.75% (2) 12.5% (3) 66.6% (4) 78.3% (5) 87.5%

Radioactivity

It is easy. But you need to understand correctly of what has been asked. As the half-life is 30 minutes, it has passed 3 half-lives after 3 hours. So, the residual fraction after the decay is $1/2^3 = 1/8$. The residual fraction after one half- life is ½. The residual fraction after another half-life is ¼. When n half- lives are passed, the leftover fraction is $1/2^n$. If you have this fact in mind, then the work is easy, So, the decayed fraction $=1-1/8 =7/8$. If we take the percentage, then 7/8 X 100% = 87.5 %. It can get wrong in between 1/8 and 7/8. If 1/8 is taken, then the answer is 12.5 %. The decayed fraction has been bold to keep you in track without going into the wrong answer. In $N = N_0 e^{-\lambda t}$ equation, N means the amount of nucleus that is left after time t. The decayed amount of nucleus is $N_0 - N$. If it is expressed as a fraction, then it is equal to $(N_0-N)/N_0$. That means $1 - N/N_0$. A similar question has been asked at the end for 6(B) in paper 2006.

18. Intensity of the noise generated by a machine is 10^{-2} Wm^{-2} By employing a noise barrier, the intensity of noise is reduced to 10^{-6} Wm^{-2} What is the reduction in the noise intensity level?

(1) 160 dB (2) 100 dB (3) 60 dB (4) 40 dB (5) 25 dB

Intensity of Sound

This is very simple as jumbo peanuts. You can get the answer by closing your eyes. There are many such old questions. The fraction when the intensity is reduced from 10^{-2} to $10^{-6} = 10^{-2}/10^{-6} = 10^4$. That means the answer is 40 dB. If you apply to the equation $\Delta\beta = 10 \log 10^{-2}/10^{-4} = 40$ dB

19. A convex lens is used to obtain a clear image of an object on a screen. The screen is located 30 cm away from the lens, and the object is at 20 cm from the lens. If the lens is now used to focus the image of a distant tree on the screen, the distance between the lens and the image of the tree is

(1) 12 cm (2) 24 cm (3) 50 cm (4) 60 cm (5) 90 cm

Refraction Through Lenses

This is simple. The question asks about the focal length of the lens. The image of a distant tree is created in the focal length. Find f using the lens formula.

$-1/30 - 1/20 = 1/f$; $1/f = -50/(-30 \times 20)$; $f = (-30 \times 20)/50 = -12$ cm.

20. **Which of the types of glass prisms shown in figure (2) can forms be used to bend a ray of light into all the given in figure (1)?**

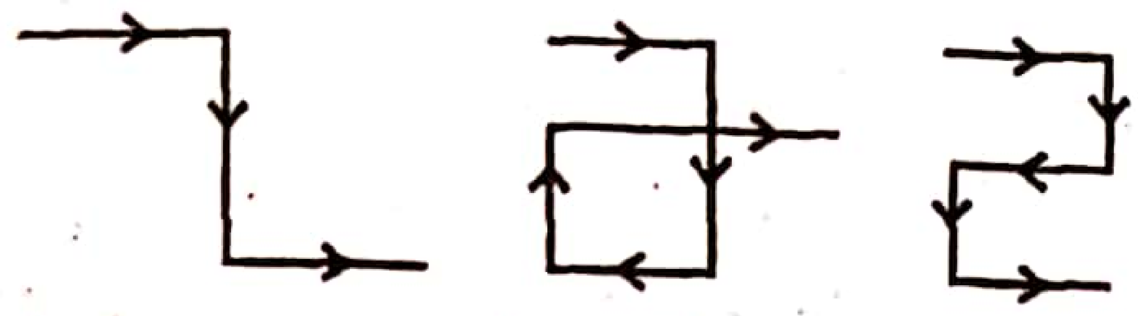


Figure (1)

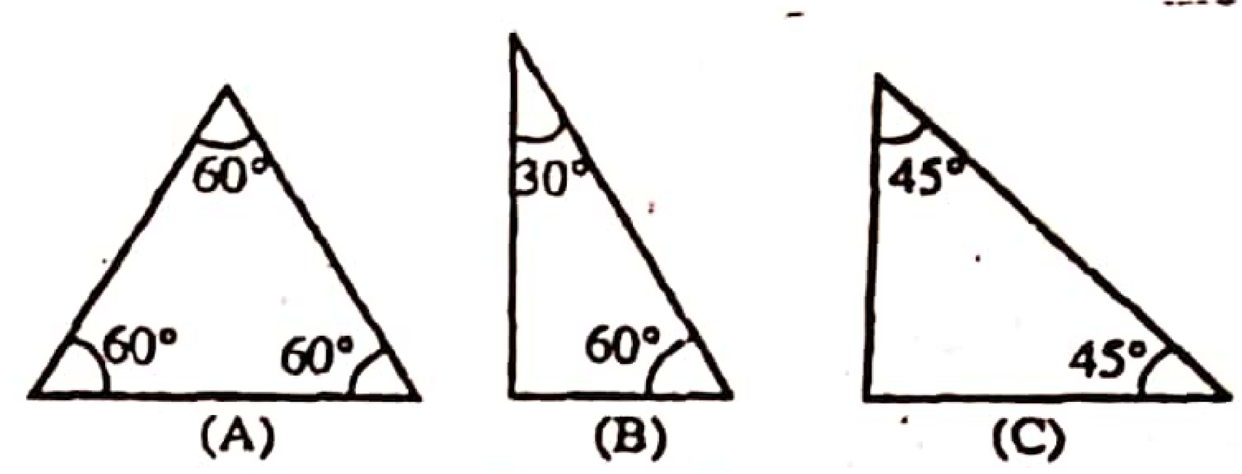


Figure (2)

(1) Type A only. (2) Type B only. (3) Type C only.
(4) Types A and C only. (5) Types B and C only.

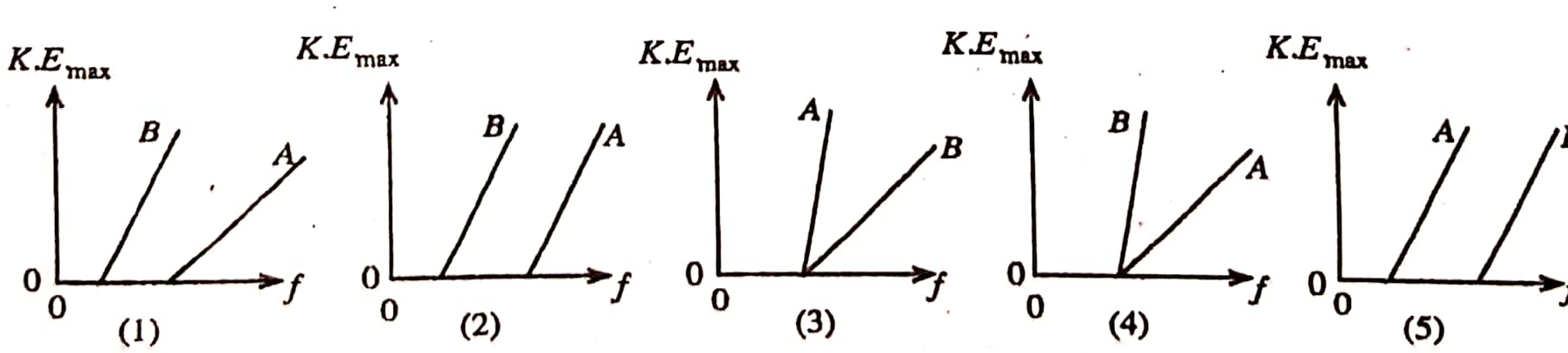
Restruction Through Prism 03

I suspect whether the children spent more time on this question. All the types of light rays that are given are turned back in 90⁰ angles. Out of them, the rays can be deflected by 90^0 only from a 45^0- 45^0 perpendicular prism. All you need to look only that fact. Have tried each prism one over the other and turned according to the given ways? Only hand and other organs will be in pain from this. Each place is turned by 90^0 angles in all the places of patterns. If the reflected ray is needed to be turned in 90^0, then the incident ray should be 45^0. There are no words about it. Can you get 45^0 incident rays from (A) and (B) prisms? The critical angle for glass - air interface is about 42^0. Therefore, if total internal reflection needs to be occurred, then the incident angle should be greater than 42^0. From 30^0 this will not occur. If it occurs after 60^0, then the angle between the incident and the reflected rays will be 120^0. Or else the ray is deflected by 60^0.

These are things that I write more. As soon as you see, you need to hit 45^0-45^0 prism. Do not try to keep the prisms and find the given patterns. Once the exam is over, if you have the desire or the fever keep (C) prisms according to the given patterns at home. However, from one prism of (C) you cannot get the patterns. In the question it has been mentioned as glass <u>prisms</u> not as a glass <u>prism</u>. Therefore, you can use any number of prisms of (C). For example, you need two prisms of (C) to get the first pattern.

Once you see such questions, you should get the clicked feeling like once you find the life partner. The people who try to keep prisms one over the other cry saying that the time is not enough. Find the partner that is matching well to yourself. The method of putting one over the other is dangerous.

21. **The work functions corresponding to two metals A and B are W1 and W2 respectively, and W1 > W2. Two surfaces . made of A and B are illuminated separately using a monochromatic beam of light of frequency. Which of the following graphs correctly represents the variation of the maximum kinetic energy (K.Emax) of the emitted photoelectrons with the frequency (f) of the incident light beam, for the surfaces made of metals A and B?**



Photoelectric Effect 11

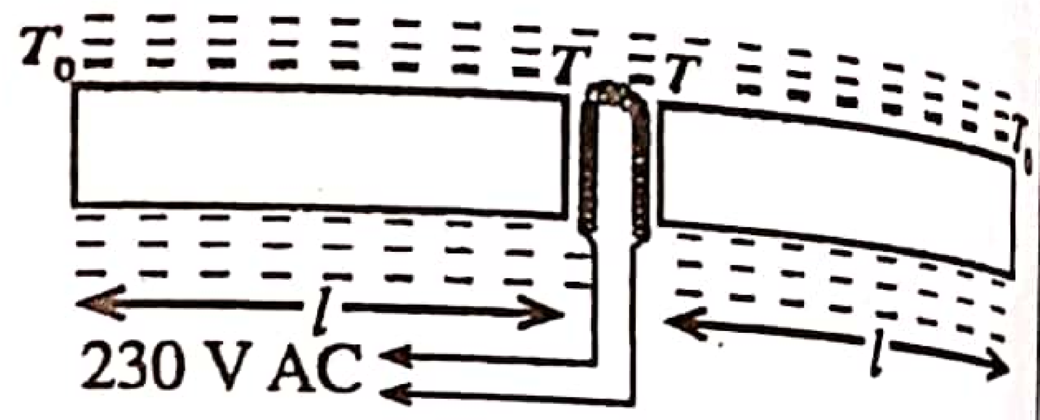
You can find the answer before reading the question. The gradient of the graph that shows the variation of f with K_{max} should be equal to Planck's constant and the negative intercept should be equal the work function. $K_{max} = hf - \varphi$. Therefore, the graphs should be parallel to each other and there should be a greater intercept in the metal that has the highest work function. All you need to look at these two facts. The correct variation is (2).

In our times, the symbol W had been used for work function. But that is not an issue for the question. Actually, does the length of the question body is more? Intelligent children do not read such questions from top to bottom. What is being asked is quickly realized by them. Middle level and bit weaker children read each word without neglecting. Afterwards, they say that the question is long and it takes time to read. Actually, what if this question is asked like this way? K_{max} and f are standard symbols.

Two metals of A and B were subjected to photoelectric effect where metal A has the highest work function than that of metal B. Which graph is correct?

I think that this is enough. This is only my opinion. Many questions can be shortened if we can have a convention of giving questions like this way. The reading time can be reduced. I believe that time has come to have a discussion regarding this matter and come to a conclusion. By this the mourning of children can be minimized.

22. Two ends of two identical metal rods of uniform cross section are placed very close to each other, and those ends are heated usins an electric heating element which supplies heat at a constant rate of P (Watts), as shown in figure. The rods are thermally well insulated as shown, and at the steady state, the temperature at free ends which are exposed to the surroundings is T_0. Assume that the entire heat energy generated by the element is absorbed equally by the two rods. If l, A and k respectively are the length, cross sectional area and the thermal conductivity of a rod, what is the temperature T of the ends of rods close to the heating element at the steady state



(1) $T = T_0 + \frac{Pl}{kA}$ (2) $T = T_0 + \frac{Pl}{2kA}$ (3) $T = T_0 + \frac{2Pl}{kA}$ (4) $T = 2T_0$ (5) $T = 2\left(T_0 + \frac{Pl}{kA}\right)$

Conductivity

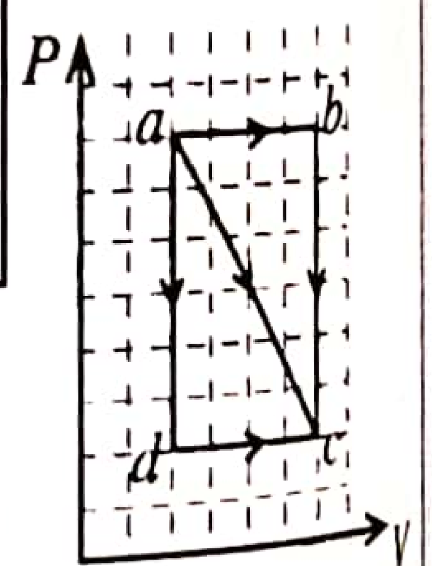
Even the question is long, you can get the answer in a very simpler way. It is a very easy question. Apply heat conductivity equation to a rod. What else to do?

$P/2 = KA\,(T - T_0)/l$; $(T - T_0) = Pl/2KA$. The answer is just in your hands. Why do they say that this question is hard? Cannot we reduce the length of this question as below?

A heat supplying device supplies heat equally to each rod with a power P. If the terms l, A and k are the length of the rod, the cross-sectional area and the heat conductivity respectively, then from what does T is given after it gets to a continuous state?

If A and k are marked in the figure, then the second sentence could have been shorter.

23. An ideal gas can expand from state a to state c along three thermodynamic paths adc, ac and abc as given in the P- V diagram. Along which of the above paths would the highest exchange of heat occur?



(1) Path adc (2) Path ac (3) Path abc
(4) Path adc and ac equally (5) Path adc and abc equally

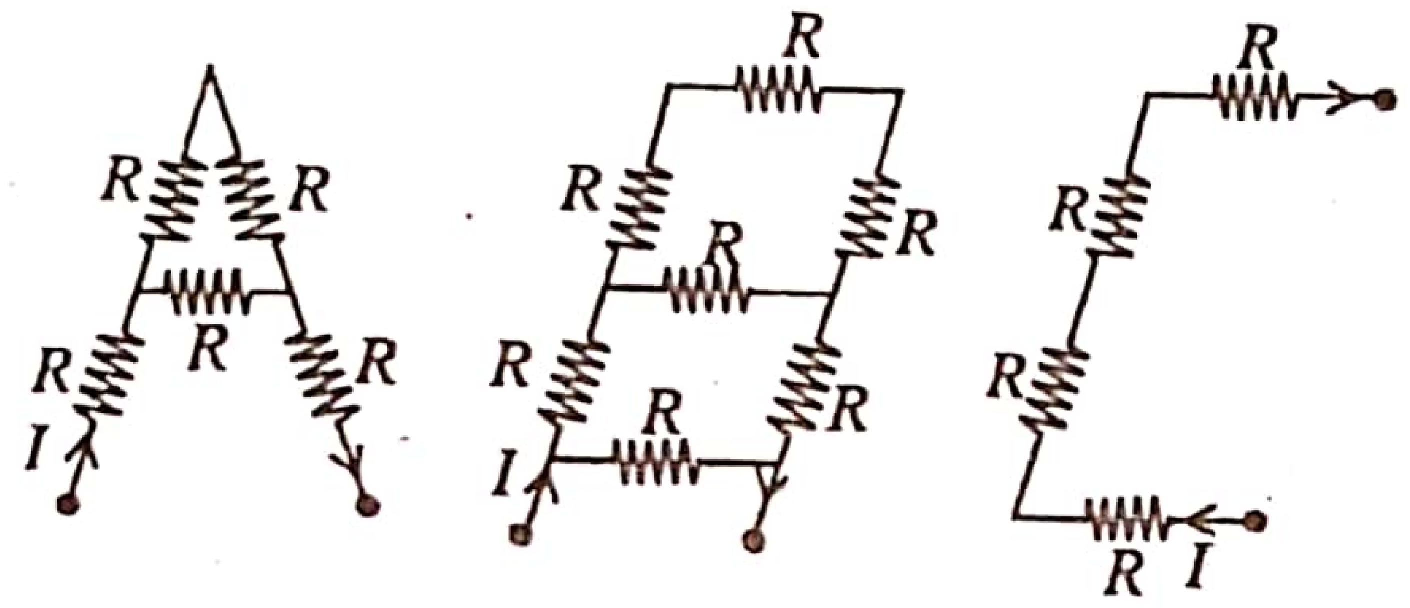
Thermadynamics

This is a nano peanuts question. A greater heat transfer occurs at the path with the highest work. ΔU is same if you go from any path. An expansion happens here. Therefore, ΔW is positive. According to $\Delta U =$

$\Delta Q - \Delta W$, if ΔU is same, then when ΔW is increased, ΔQ also should be increased. Otherwise, you cannot keep ΔU at a constant value. You can see that the highest work is done through the path of abc if you have even one eye. The greatest area is made with the volume axis from the path that goes from a to b. There is no doubt that you have done many of such questions. Only about heat transfer is being asked. There is no need find whether it is positive or negative.

If needed, we can do the analysis further. When considering a and b points, the temperature of the gas in b should be higher than the value at a. By applying $PV = nRT$ separately to a and b points you can come into this conclusion. P is same but V increases. Therefore, ΔU should be positive in a→ b path. In the path of a→ b, ΔW is also positive. Therefore, according to $\Delta U = \Delta Q - \Delta W$, ΔQ should also be positive if ΔU needs to be positive. It should be also $\Delta Q > \Delta W$. Analyze like this if you take a question.

24.



(A) (B) (C)

The same current I is sent through resistor networks A, B and C as shown in above figure. If all the resistors in the networks are of equal magnitude, the maximum power is consumed by

(1) the network A.
(2) the network B.
(3) the network C .
(4) the networks A and B equally.
(5) the networks B and C equally

Heating Effect of Electric Current **08**

Were all tempted to get expressions for this question? There is no need. You do not need it. Same current goes across the networks. Therefore, the power consumption can be obtained by I^2R. The equivalent resistance is maximum in the circuit of (C). Cannot you just see that? All the resistors are being connected in series there. In the other two, all the resistors are not being connected like this way. There is nobody to compete with the maximum equivalent resistance of 4R. If you can find a resistor arrangement greater than 4R, then you will be able to get the Nobel prize. So, why do you look at the networks of (A) and (B)? Is not the answer (3)?

Can you compare the power consumptions? Can you get that (B) circuit has the least resistance by logic? I will give a hint. When you come from top to bottom, R at the bottom is in parallel with another resistance at the end. Therefore, the equivalent resistance should be lesser than R value. The equivalent resistance of (A) should be greater than some value of 2R. But it cannot be 3R. My God! Try to argue like this way. If my older brain can do it, then why cannot your younger brains do it? The problem is the laziness to argue. You would like to look at gossips in Facebook than this method.

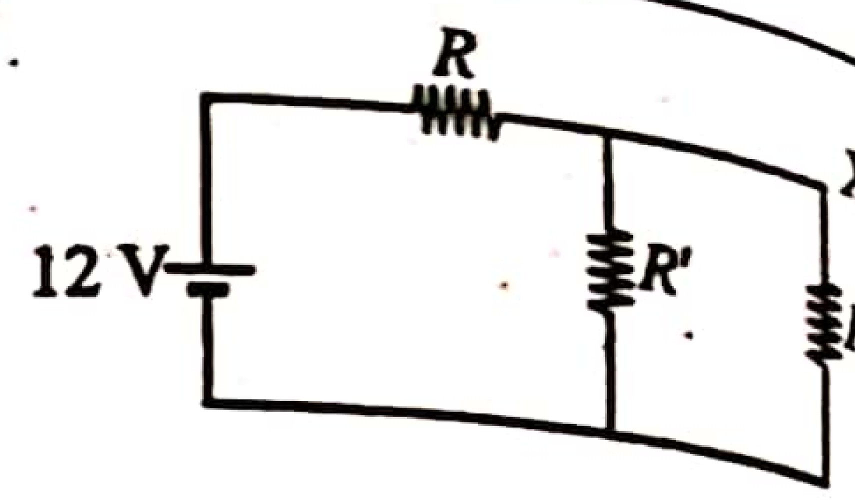
25. A 5W electronic device having a resistance of 5 Q is operated by receiving power from a 230 V main supply through a transformer. The ratio, $\dfrac{\text{Number of turns in the primary coil}}{\text{Number of turns in the secondary coil}}$ of the transformer is

(1) 46 (2) 23 (3) $\frac{10}{23}$ (4) $\frac{1}{23}$ (5) $\frac{1}{46}$

Mutual Induction **08**

This is just a problem. Find the voltage for the device. $W = V^2/R$, $5 = V^2/5$, $V = 5V$; $N_P/N_S = V_P/V_S = 230/5 = 46$. You can just get that this is a step-down transformer. Then you can remove the choices of (3), (4) and (5). The number of turns in the primary should be greater than the secondary.

26. **In the circuit shown, the voltage at X is found to increase by 4 V when R' is removed. The resistance of R' is equal to**

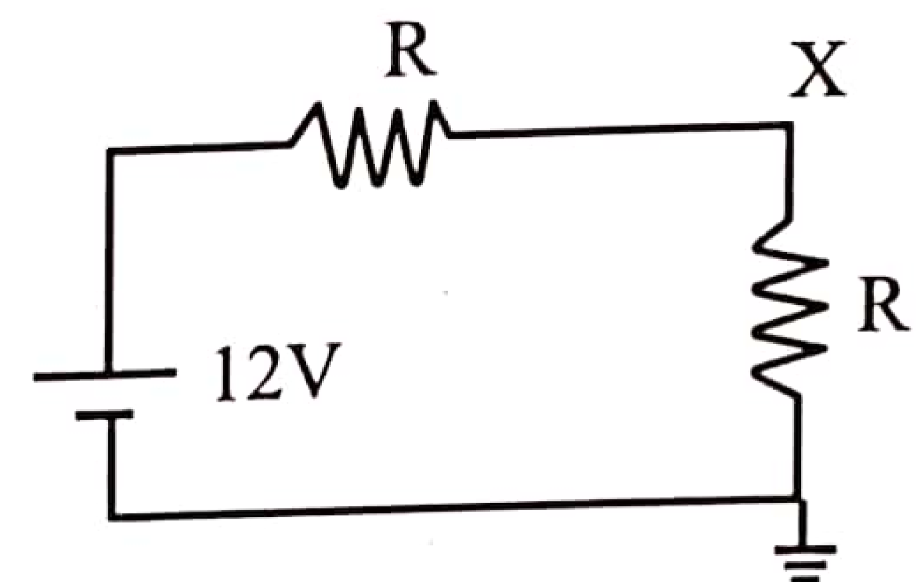


(1) $4R$　　(2) R　　(3) $\frac{R}{2}$

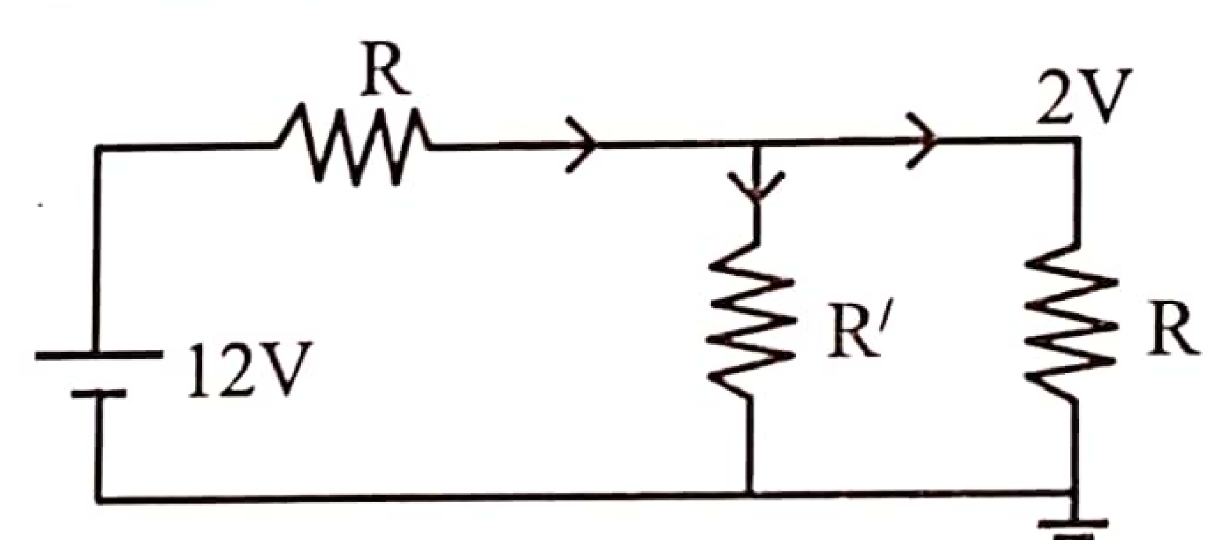
(4) $\frac{R}{4}$　　(5) $\frac{R}{6}$

Ohm's Law Combinations of Resistance

The answer can be easily obtained if it is solved form the logic. As the voltage of point X is mentioned, the negative end of the cell should be earthed. Even it is not essential, then the logic is easy. When R' is removed the circuit will look like this way.



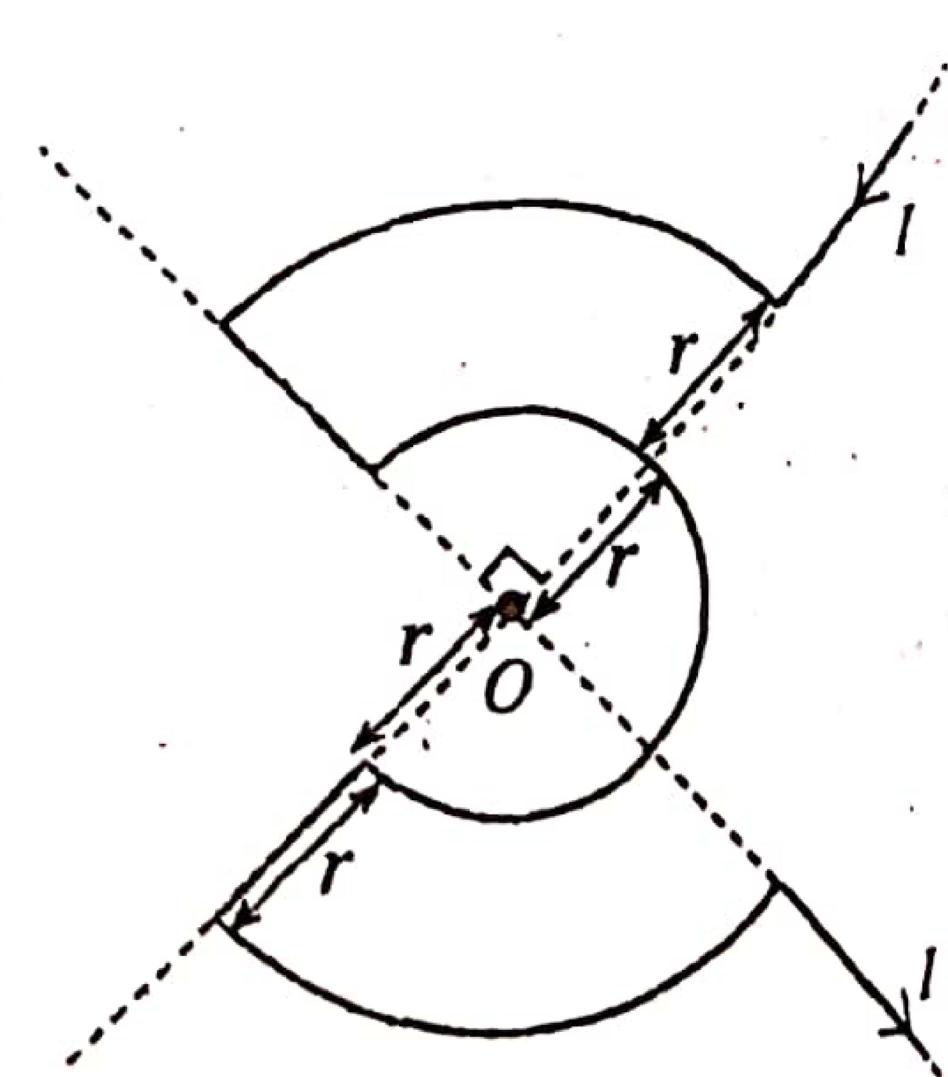
Then the voltage of point X should be 6V. Why? 12V is divided equally with two Rs. When R' is removed, then the voltage of X is increased by 4 V, then when R' is there, the voltage of X should be 2 V (2+ 4 = 6). Now go to the previous circuit. You do not have to draw the above circuit even if I drew. You can get the voltage of X as 2 V by the memory without drawing the circuits. Now put 2V near X in the paper. Cannot you find R' now? This can be found from two methods. If the voltage is 2 V across the equivalent of R and R', then voltage is 10 V across the rest of R in the circuit. Now if you consider the total of the currents,



10/R = 2/R + 2/R'; 1/R' = 5/R-1/R; R' = R/4

The other way is that the equivalent resistance should be R/5 across the parallel equivalent of R and R' as the voltage across the equivalent of R and R' is 2V and as the voltage is 10 V across the rest of R on the left side. If 10 is across R, then across R/5 it will be 2. Now for the parallel equivalent circuit, 5/R = 1/R + 1/R' where R' = R/4.

27. **A piece of wire is bent into the form shown in figure and a current of I is passed in the direction shown. The magnitude of the magnetic flux density at the point O is**



(1) $\frac{\mu_0 I}{4r}$　　(2) $\frac{\mu_0 I}{8r}$　　(3) $\frac{3\mu_0 I}{2r}$

(4) $\frac{\mu_0 I}{2r}$　　(5) $\frac{3\mu_0 I}{8r}$

Magnetic Effects of Electric Current

There are many such arrangements in previous papers. Mark the direction of the first current on the wire. Keep your rough work to a minimum. The current flows in clockwise direction in the smaller circle. In addition, there ¾ of a full circle in the small circle. Therefore, the magnetic flux density from it is $\frac{3}{4}\frac{\mu_o I}{2r}$. The biggest circle is half of a complete circle and the current flows in anti-clockwise direction there. The value of B from it is $\frac{1}{2}\frac{\mu_o I}{2(2r)}$. B from the smaller circle is towards the paper whereas B from the bigger circle part is out of the paper. Also, B value from the smaller circle is bigger in magnitude. It is also closer to the centre. Even it is with many parts. Therefore, the net magnetic flux density at O is $\frac{3}{4}\frac{\mu_o I}{2r} - \frac{1}{2}\frac{\mu_o I}{4r}$. You only have to write this on your rough paper. Simplify it afterwards $\frac{2\mu_o I}{8r} = \frac{\mu_o I}{4r}$. As there is common denominator in both of the terms, it is easy to simplify.

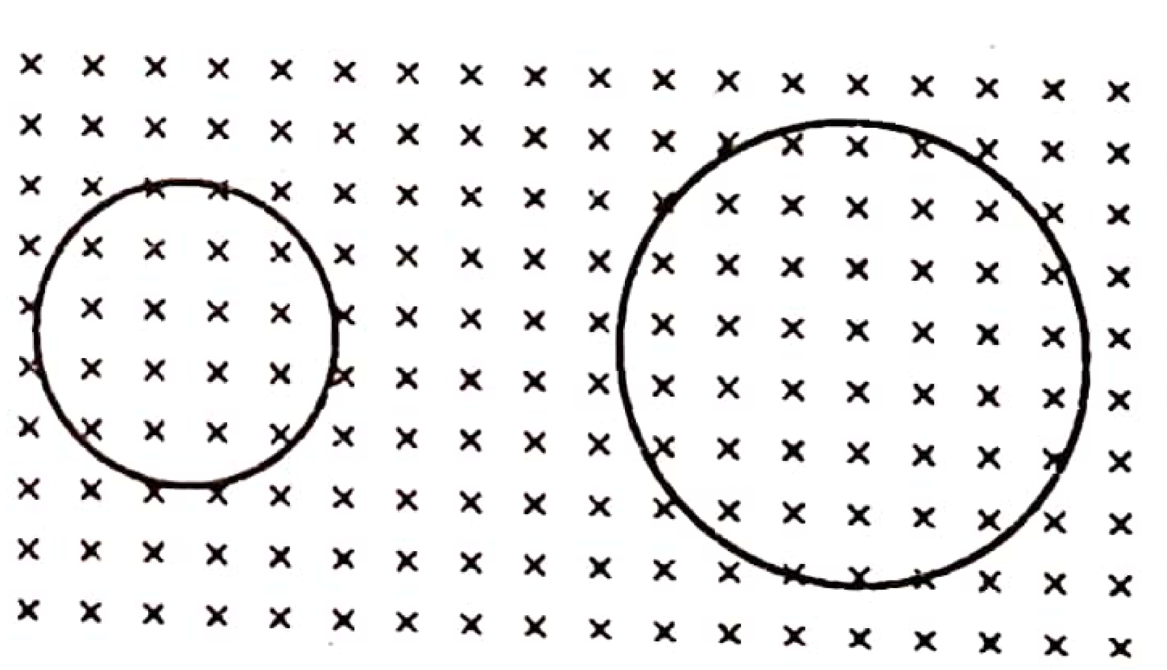
28. Two identical strings are separately subjected to a tension T. When plucked at the middle, each string I produces waves of frequency f. Now, if the tension of only one string is reduced to 0.81T and the strings I are plucked at the middle simultaneously, five beats can be heard during one second. The value off is I

(1) 25 Hz (2) 50 Hz (3) 75 Hz
(4) 90 Hz (5) 100 Hz

Transverse Waves **03**

The length of the strings is not changed. In both occasions, as the string has been vibrated from the middle, the generated wavelengths are not changed. Therefore, frequency f, is proportional to the sound speed of transverse waves (v) $v \propto \sqrt{T}$. That means ; $f \propto \sqrt{T}$. The square root of 0.81 is 0.9. Beat frequency = f - 0.9 f =5. Then f =50 Hz. This is the method with least equations and mathematics. When proportionality constants are being inserted, it will be lengthy. If you try to do it in the lengthy way, then $f = k\sqrt{T}, f_1 = k\sqrt{0.81T} = k\ 0.9\sqrt{T}$. Therefore, $k\sqrt{T} - 0.9\ k\sqrt{T} = 5;\ k\sqrt{T} = 50 = f$. You should see that 0.81 is given for the purpose of finding the square root conveniently.

29. An electron and a proton travel with equal speeds around two circular paths shown in the diagram (drawn not to scale) under the influence of a uniform magnetic field. If the direction of magnetic field is perpendicular and into the plane of the paper,



(1) the electron travels clockwise around the small circular path and the proton travels counter-clockwise around the large circular path.

(2) the electron travels counter-clockwise around the small circular path artd the proton travels clockwise around the large circular path.

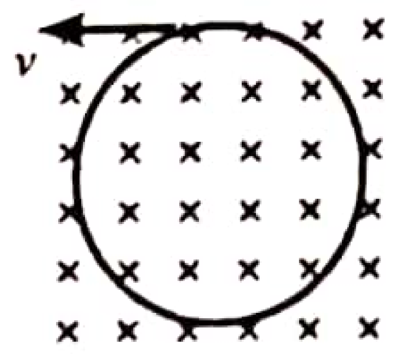
(3) the electron travels clockwise around the large circular path and the proton travels counter-clockwise around the small circular path.

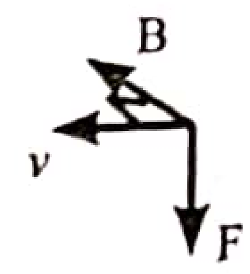
(4) the electron travels counter-clockwise around the large circular path and the proton travels clockwise l around the small circular path.

(5) the electron travels counter-clockwise around the small circular path and the proton travels counter-clockwise around the large circular path.

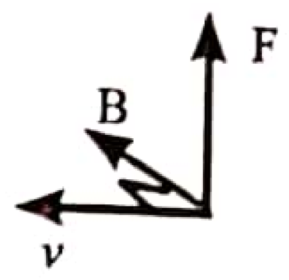
Force on a Moving Charge in a Magnetic Field **07**

It is a very easy problem. You need to think only about the moving direction. According to qvB = mv²/r, the proton with the higher mass should travel in a larger circular path. As q and v are equal (in magnitude), $r \propto m$. To find the travelling direction, you do not need to consider an instance where an electron is entering a magnetic field with v speed. You do not have to think whether the particles are coming from top or bottom or from the side. In the question it has been given that it is going in a locus. Therefore, what you need to do is to decide whether it is clockwise or anticlockwise and check whether that decision is correct or not. Let us consider the electron. Consider that it is going anticlockwise and mark the direction of v.





If a positive charge

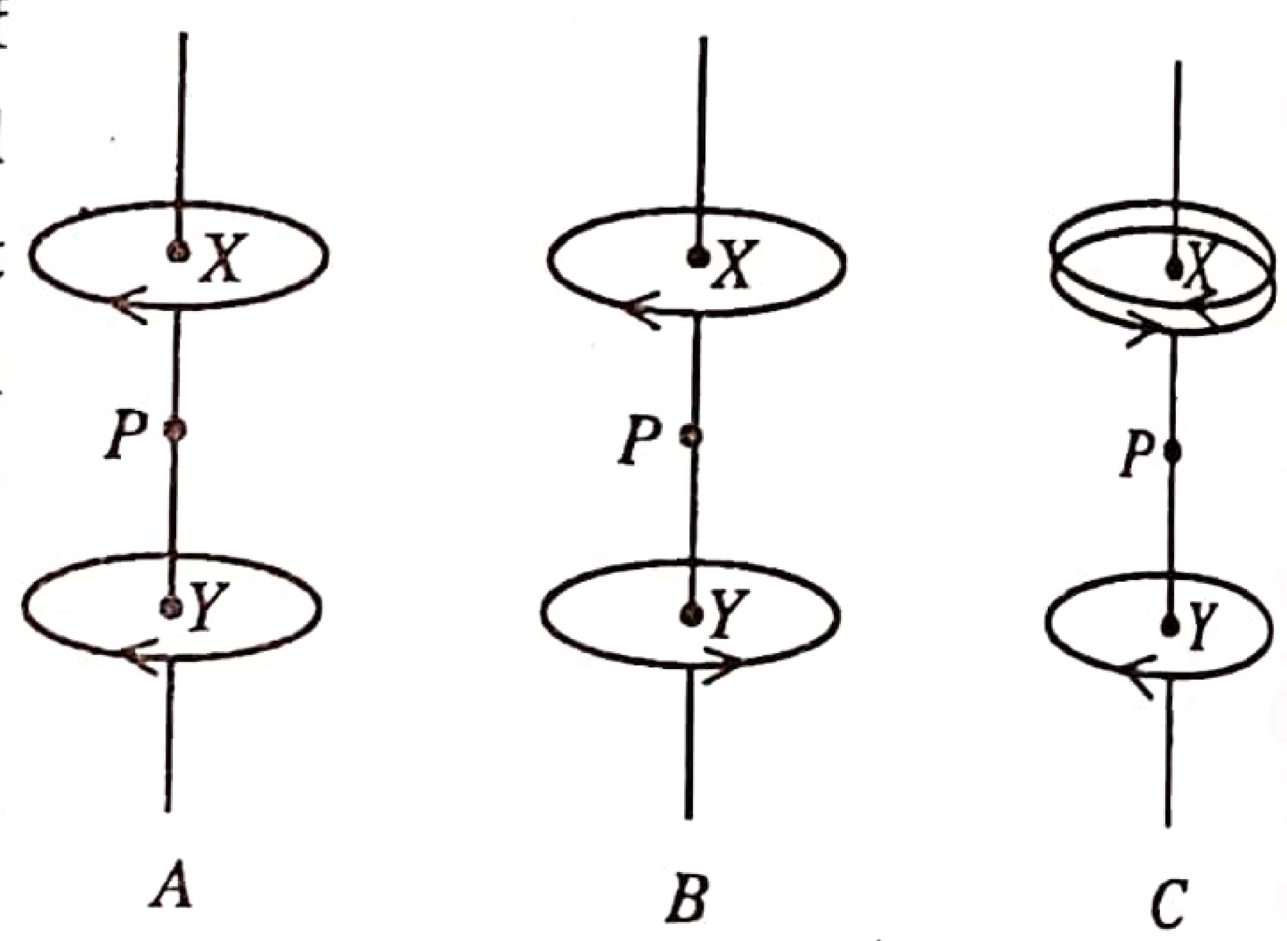


If a negative charge

Next, decide whether the chosen direction of force of qvB is correct. It seems that the decision is not correct. If the electron is going anticlockwise, then the force F should be downwards (towards the centre). But the force F is towards up. Therefore, there are no two words about it. The locus should be in clockwise direction. If we decide that the direction is clockwise at the beginning, then it seems to be correct.

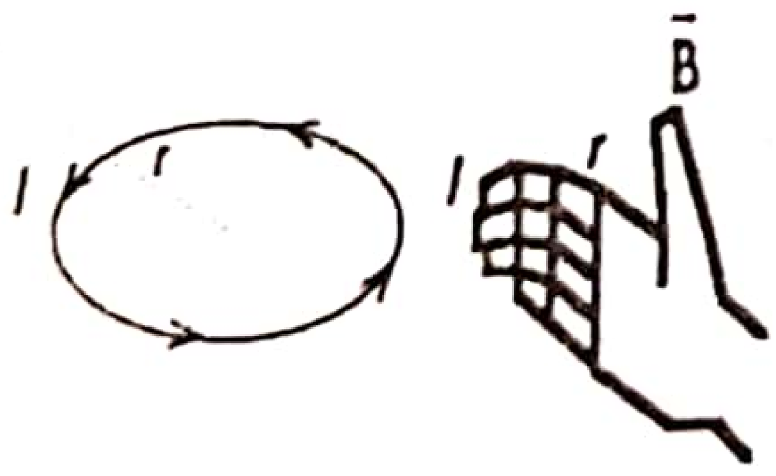
Therefore, the electron definitely travels in a clockwise locus. If the electron travels in a clockwise direction, the proton should go in an anticlockwise direction. Why? Because the charge of the proton is positive. You do not have to think something new about it. If the electron is clockwise, then the proton is anticlockwise. As the mass of the proton is higher than the mass of electron, the radius of the locus of the proton should be greater.

30. Identical loops in the three arrangements A, B and C of circular loops centred around vertical axes, carry equal currents in the directions shown in figure. In the arrangement C there are two separate loops very close to each other with a common centre at X. In all three arrangements the loops are separated by the same distance XY and P is the mid-point of XY. If the magnitudes of the magnetic flux densities at P in the arrangements A, B and Care BA, BB and BC' respectively, then



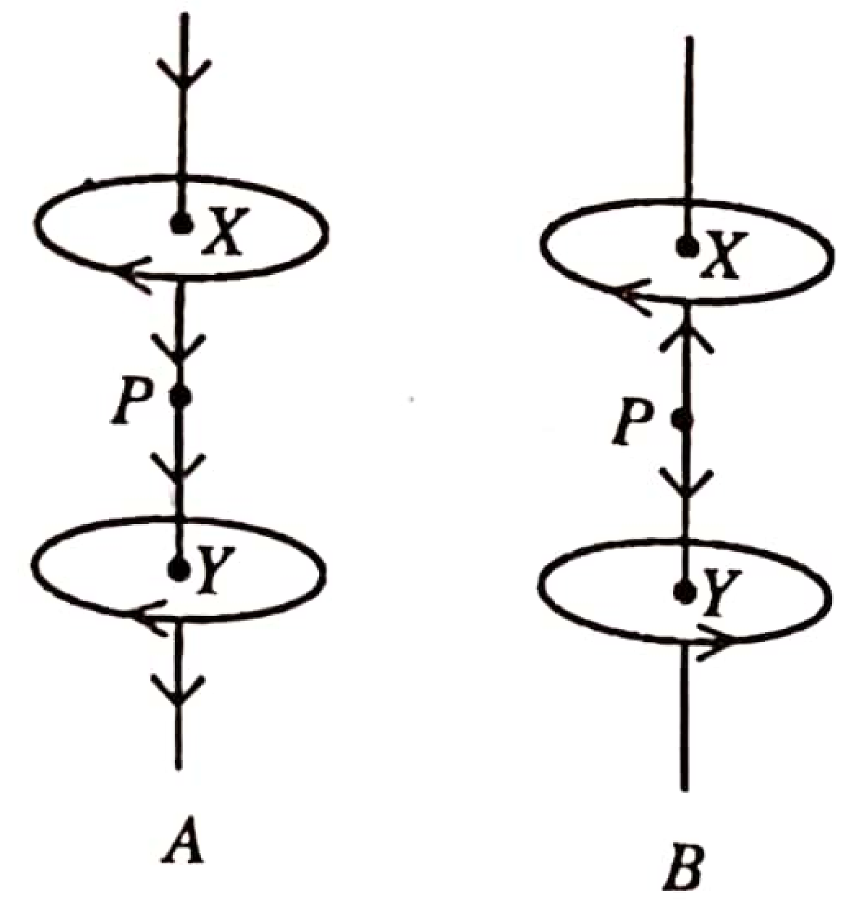
(1) BA > BB > Be (2) BA > Be > BB (3) B8 > Be > BA

(4) Be > BB > BA (5) Be > BA > BB

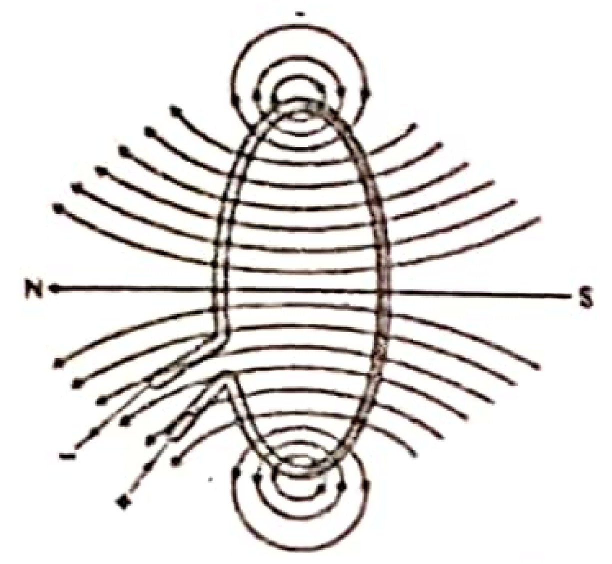


Writing an expression of magnetic flux density for a point that is not in the axis of the centre of a circular loop carrying current is not in the syllabus. But simply its direction can be obtained. Keep the right-hand thumb perpendicular to the other fingers and direct the finger tips to the current carrying direction. That means direct the finger tips to the direction of the current and place the fingers on the loop. Then direction of the magnetic field is obtained by the direction that the thumb is pointing. Look at the figure.

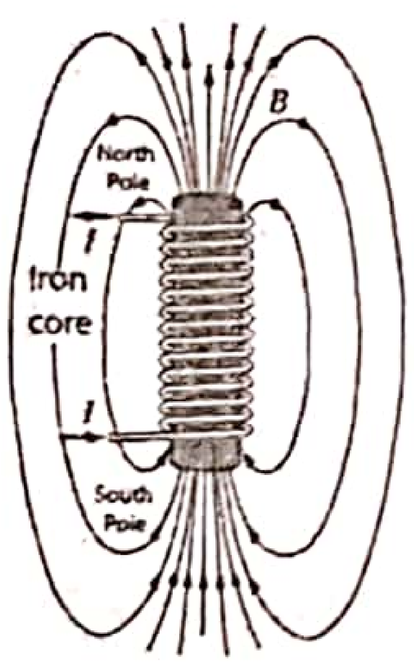
According to this, you can quickly see that the magnetic flux densities are added at point P in arrangement A whereas in arrangement B of point P, the magnetic flux densities are towards opposite direction.

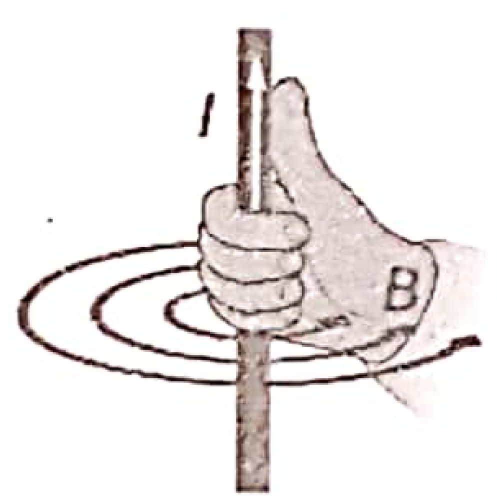


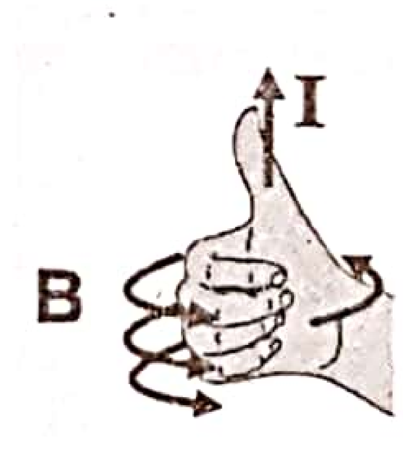
Actually, the magnetic flux density is zero in P of the arrangement of B. It is a null point. In the arrangement of C, as two loops carrying equal currents to opposite directions are placed near to X, the magnetic flux density created at P which is in equal distance will be cancelled off. Therefore, the magnetic flux density of P in arrangement C is given by the loop with single Y as the centre.

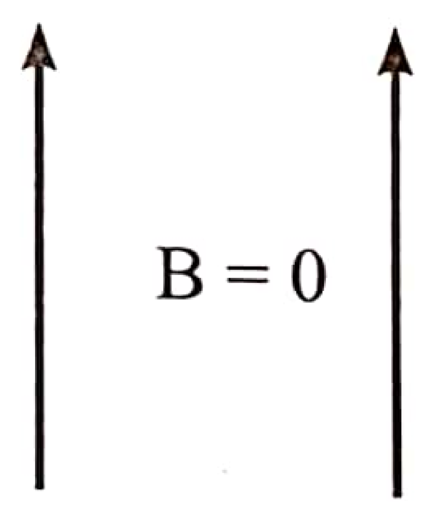


Therefore, if you think in a simple way, if B is the magnitude of the magnetic flux density at P from one current loop, then for arrangement A, $B_A = 2B$; for arrangement B, $B_B = 0$ and for arrangement C, $B_C = B$. Now quickly you can decide that $B_A > B_C > B_B$. Do not try to equal the straight wires which carry current with current carrying loops. The net magnetic flux density in the middle is zero in two parallel wires which carry an equal amount of current to the same direction. If the thumb is directed to the direction of the current, then the direction of the magnetic field is obtained from the direction of the fingertips of the other fingers. According to that, B = 0 in the exact middle.

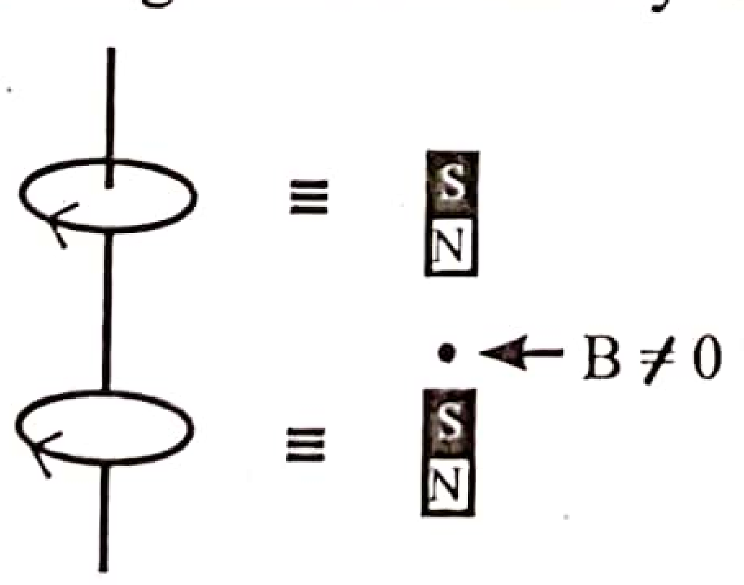


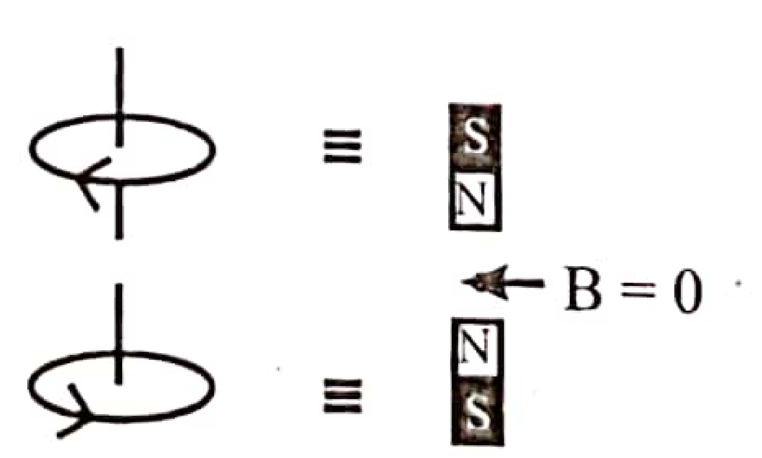






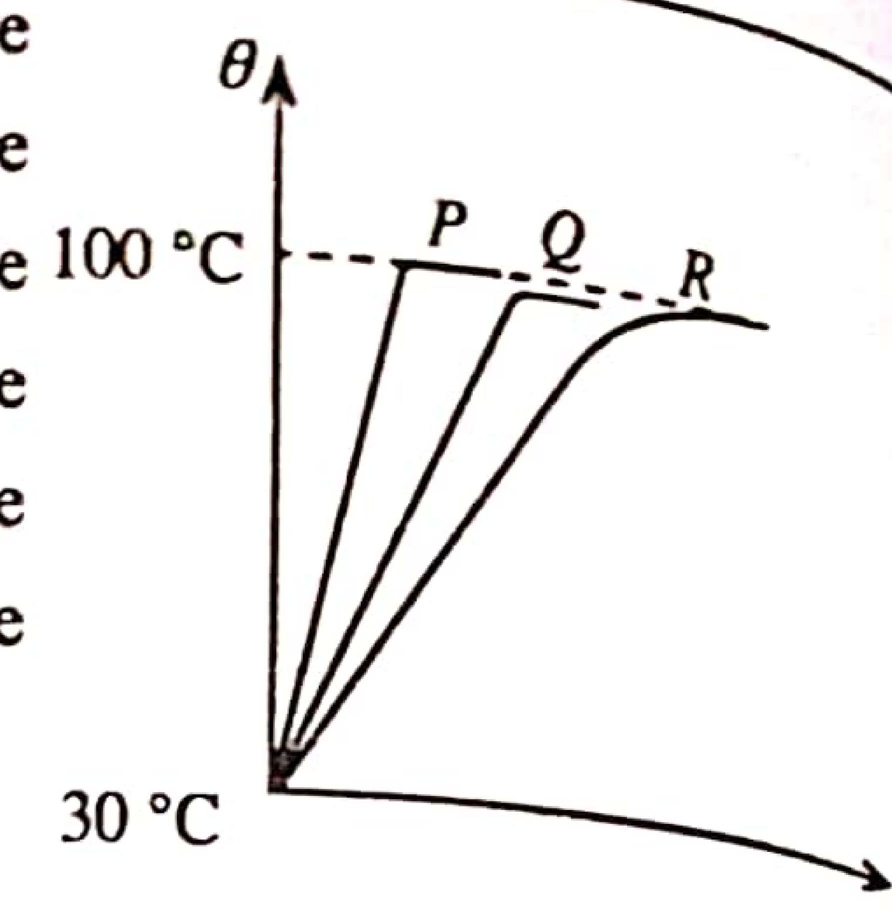
B of the middle of two parallel circular loops which carry equal current to the same direction will not be zero. Here B is acted upon the same direction and the net gets doubled. So, do not try to get confused with these two arrangements. Current carrying circular loops which are closer together which carry current to the same direction are like a solenoid when taken as a whole. When the turns of the solenoid increase, the magnitude of the magnetic flux density is increased across the axis of the solenoid. It will never be zero.





In anohter way, a current carrying loop can be considered as a small bar magnet. If you direct the fingers to the directions as mentioned before, the thumb is pointed towards the direction of the noth pole.

31. Three different types of thermometers, P, Q and R having a temperature range of O- 110°C, and kept at room temperature of 30°C were simultaneously dipped into a large oil bath, maintained at 100 °C at time t= 0, and their readings (0) were recorded with time(t). Curves in figure show the variation of O with t for three thermometers. Consider the following conclusions made about the thermometers after analyzing the three curves.



(A) P is the most sensitive thermometer.

(B) Thermometers P and R are accurate but not Q. (C) The scale of thermometer R is not linear

Of the above conclusions,

(1) only A is true. (2) only B is true. (3) only A and B are true.

(4) only B and C are true. (5) All A, B and C are true

Thermometry

This is a question that can get wrong. There are Physics and logic in this question. Such type of questions has been checked several times before. The secret of these questions is getting the true/false nature of the sentences based only on the given facts. What can we decide from the variation of P? The temperature has risen very quickly. The relevant characteristic for this is the quick response. Can you decide that it is sensitive as it is quick in response? Sensitivity is the change of a thermometric characteristic in a large amount for a certain temperature change. Then even a small temperature change can be measured.

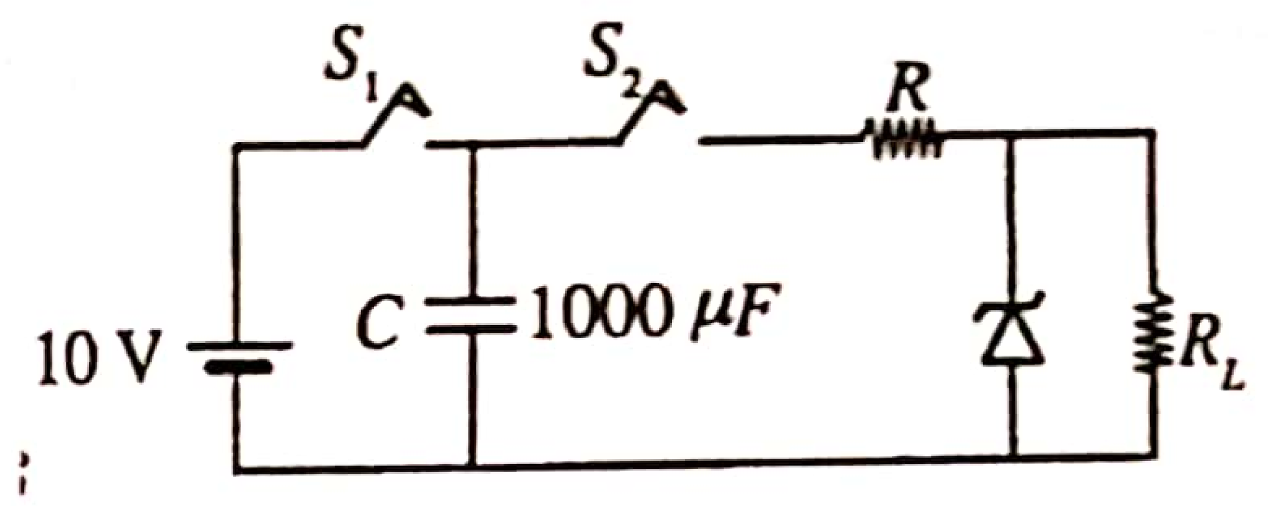
So, you cannot say that (A) is correct exactly. It can be true but we don't know. The given variation indicates that this thermometer is a quickly responding thermometer. People with quick responses cannot be sensitive. Do not try to lose your sensitivity while trying respond quickly. Then your partner will leave you. The same characteristic is not represented from these two characteristics.

It can be understood that (B) is true. The container has oil in 100°C. Therefore, the final thermometric reading should be 100°C. Finally, θ does not show 100°C. Therefore, it can be branded as an inaccurate thermometer. If this is a mercury thermometer, then the mercury column is stuck at 98°C. It can be argued that R thermometer is not accurate. To be accurate, it is not a compulsory factor that the thermometric characteristic should vary linearly. It is sufficient if it is changing. You can calibrate if it does not vary linearly.

Sentence (C) can be correct. But we cannot say accurately. Here it has been mentioned that the scale is not linear. That means at higher temperatures (when it nears to 100°C temperatures), the markers are put in a gradually increasing way for the distance between two adjacent scales. If so, the given variation is obtained for R. But is this the only reason for this variation? Cannot this be a reason where the thermometric characteristic is no liner for higher temperatures? Actually, it can be a reason.

There can be a higher probability that the thermometric characteristic is not linear instead of the scale is being non-linear. That means with the temperature increment, the rate of increment of the thermometric characteristic has been reduced. Therefore, we cannot decide that (C) is true. Only (B) can be decided as true.

32. Breakdown voltage of the zener diode in the circuit shown is 5 v R_1 is a suitable resistor. The capacitor C is first charged to 10 V by closing the switch S_1 and opening the switch Subsequently, S_1 is opened and S_2 is closed. Consider the following statements made about the functioning of the circuit after S_2 Is closed.



(A) Voltage across R_1, will be 5 V so long as the capacitor voltage is adequately above 5 V.

(B) Time period through which the voltage across R_1 remains constant does not depend on the value of the capacitance.

(C) Potential drop across R gradually decreases with time.

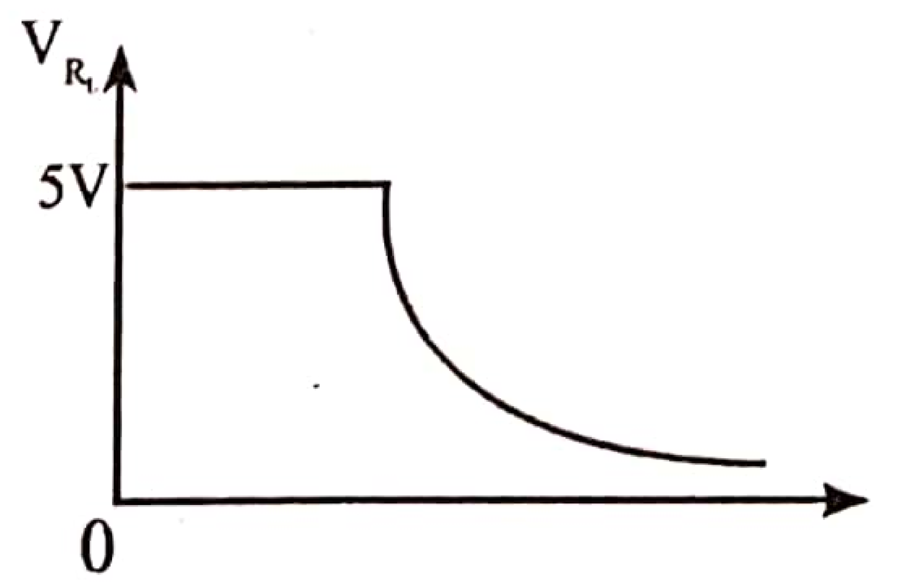
Of the above statements,

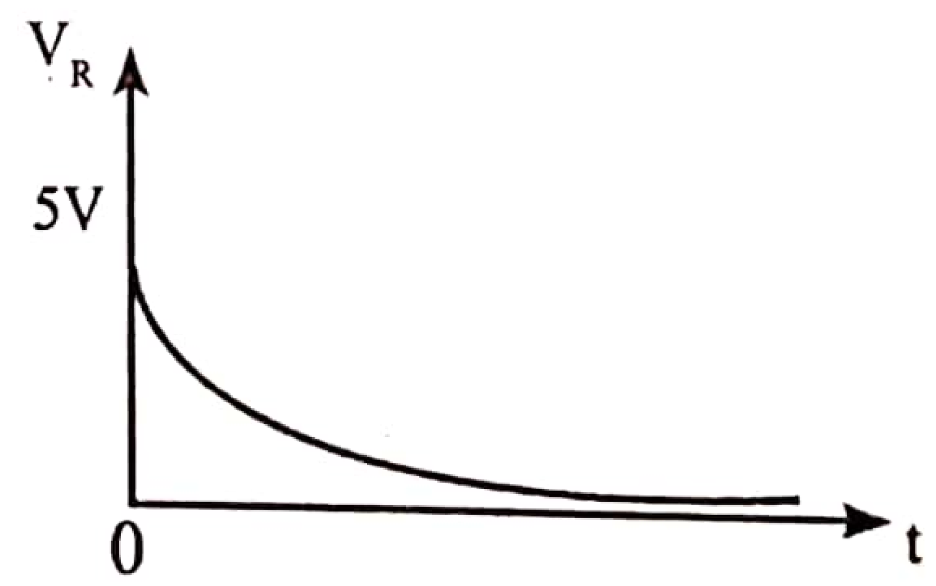
(1) only A is true. (2) only C is true. (3) only A and B are true.
(4) only A and C are true. (5) All A. B and C are true.

Does not this circuit part is equivalent to the circuit part that voltage is controlled in the bridge rectifier with the smoothing capacitor? If you can remember 9(B) question of paper 2013, then it will be easy to get the answer. As you read, you get to know that statement (A) is true. The breakdown voltage of the Zener diode is 5 V. Therefore, to keep 5 V across the voltage across R_L or the voltage the Zener diode, the voltage across the capacitor should be kept higher than 5 V. If the giving person does not have an excess, then how can the customer have more? What is meant by as long as it is higher enough because the part of voltage across the capacitor drops across R. R is the protective resistance that is used to protect the Zener diode.

In paper 2013, such a question has been asked. What is the advantage of using a higher value than a lesser value for the smoothing capacitor? You can understand that sentence (B) is wrong according to the relevant answer for this question. If C is lesser, then the capacitor quickly discharges. If so, the voltage across the capacitor quickly drops. Then the time duration that the voltage is kept at 5 V across the Zener diode at a sufficiently higher value gets reduced. If the discharge can properly be on hold, then the time duration that the voltage of R_L remains constant gets increased.

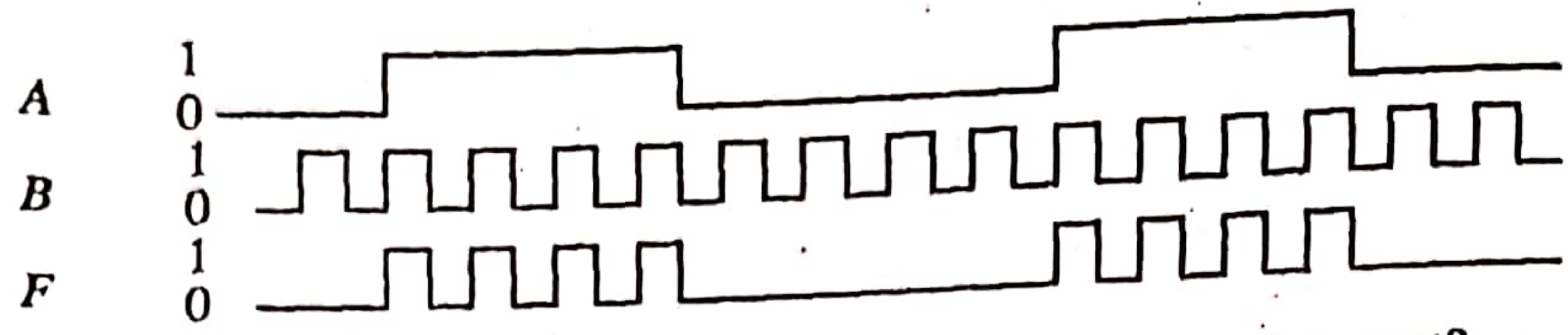
It can be realized that (C) is just true. The voltage drop across R cannot be increased with the time. However, the capacitor discharges across R. As the rectification circuit, the capacitor does not get recharged again and again. When a charged capacitor is connected across a resistor, the charge discharge of the capacitor cannot be stopped. If the road is made to go and there is nobody to hold back, then it will be gone. If needed, the variation of time (t) across R_L and R can be plotted.



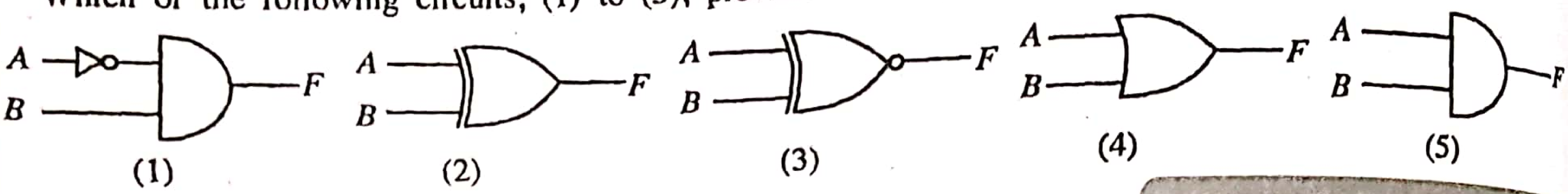


When t = 0, V_C = 10 V and V_{RL} = 5V. Therefore, V_R = 5V. With time V_C and V_R values get reduced gradually. V_{RL} is at a constant value for a certain time duration [$V_C = V_{RL} + V_R$].

33. A and B shown below represent the logical inputs applied to circuits (1) to (5) given below, and F represents the expected output from the circuit.



Which of the following circuits, (1) to (5), provides the expected output?



Logic Gates

There can be different ways that this question can be solved. I will do like this. If you can look at the input and the output and get the truth table, then the answer is in your hand. In the rough sheet, write down A, B and F and take your eyes from left to right across each input and respective output. First you will get A = 0, B = 0, F = 0. Look at the next set. That set has A = 0, B =1, F = 0. Next you will get A = B= 0, F =0. No use. Do not write. Next, you will get A = B = 1, F = 1. If you go like this way, the truth table will be completed. Is not this a AND gate? Do you need to look far more?

A	B	F
0	0	0
0	1	0
1	1	1
1	0	0

34. Which of the following is not true regarding an npn transistor and an n-channel junction field effect transistor (JFET)?

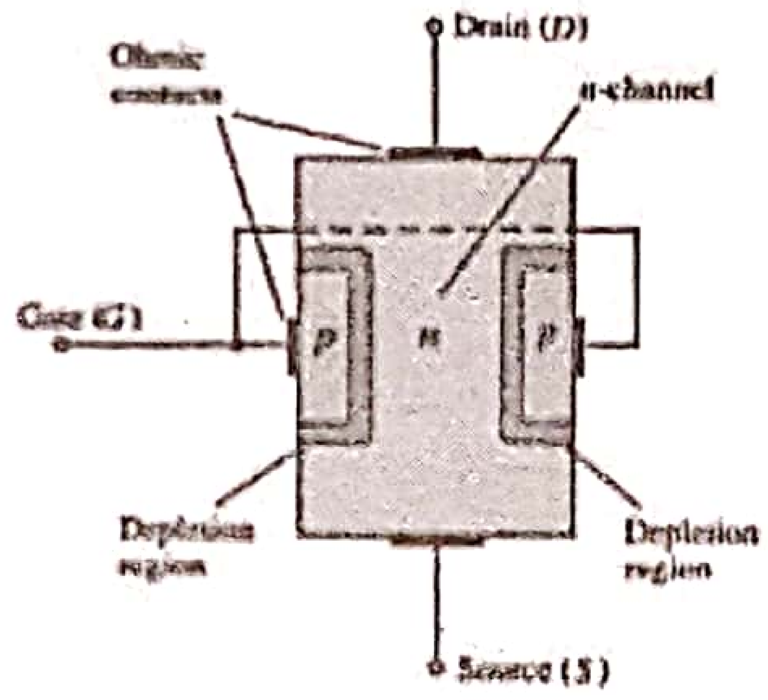
	npn transistor	*n*-channel JFET
(1)	Has two *pn* junctions.	Has only one *pn* junction.
(2)	Base-emitter junction is forward biased when operating in the active mode.	Gate-source junction is reverse biased during the operation.
(3)	An arrow is marked on the emitter of the transistor symbol.	An arrow is marked on the source of the transistor symbol.
(4)	Both free electrons and holes participate in the operation of the transistor.	Only free electrons participate in the operation.
(5)	Magnitude of the current through the collector depends on the base-emitter voltage.	Magnitude of the current through the channel depends on the gate-source voltage.

Transistors

As this is the first time that a question is given under field effect transistor (except the model paper), I feel it is better to write a description about this.

Before the semiconductor devices, the signals are amplified by using the vacuum tubes. As the vacuum tube technology is not in the present usage, nobody teaches about it. We were taught about vacuum tubes when we were at the university. Even the radios of that time were run by vacuum tubes/ valves. Those radios were big in size like big boxes (Pakis boxes). Everything got smaller when the semiconductor devices like transistors came into usage. It is profitable for the pocket to use semiconductor devices from the energy consumption side. From 1945s, Shockley (one of the scientists who won the Nobel prize in 1956 and assisted to discover the transistor) wanted to use semiconductor and make an amplifier device. He argued that a current flow across a semiconductor should be able to be controlled by an electric field. The term 'field effect' is also used due to this reason. Actually, he wanted to replace the vacuum triode tubes that was used in the amplifier circuits at that time with a semiconductor device. Even his initial attempt was fruitless, after 20 years the Bell Labs of America made the first FET. Even if you do not get successful at the beginning, never give up. Even the physical process of current flow in FET is completely different from the vacuum tube, The obtained characteristic is identical. This will be clear to you by the end of the explanation.

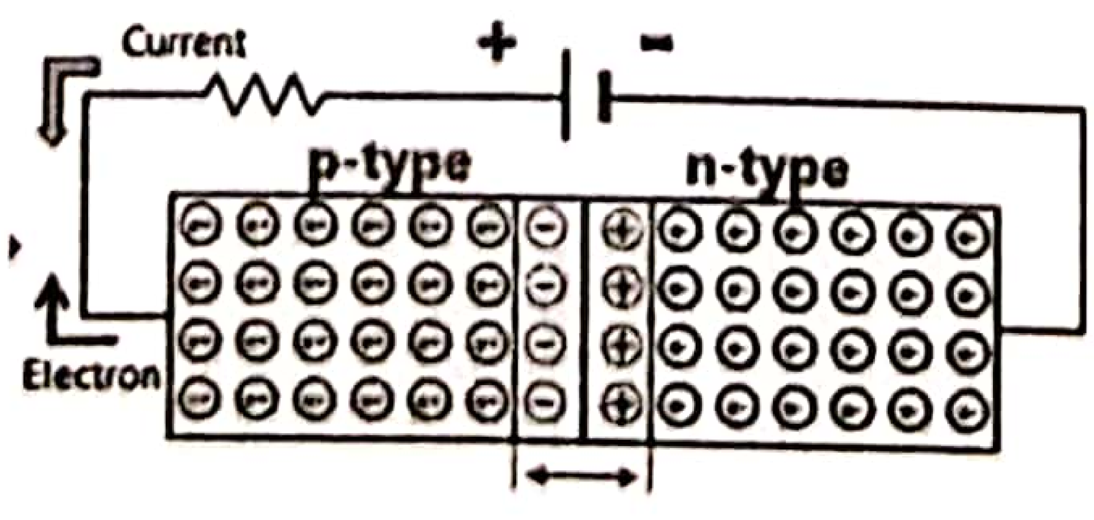
Now let us look at the functionality of a JFET (Junction Field Effect Transistor). There is a semiconductor piece of n type known as n channel transistor in the middle which is heavily doped with a p type semiconductor piece. Look at the figure.



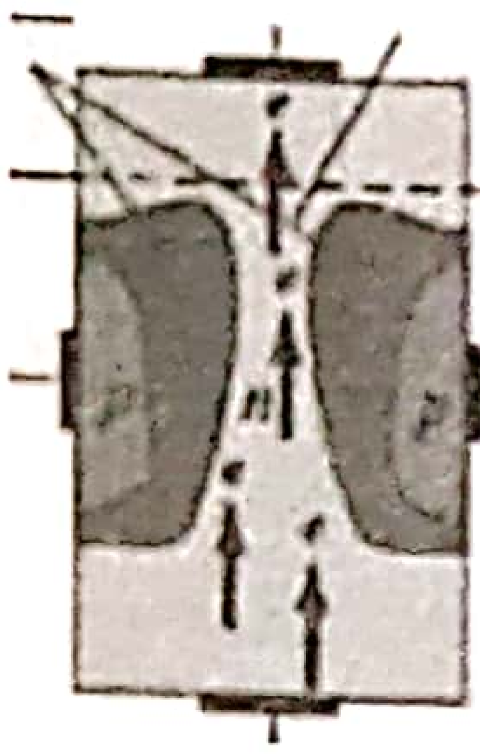
Around the n type semiconductor piece, p type semiconductor piece is there as a ring. Therefore, here you have a pn junction. When the figure is drawn in two dimensions, it is seen as two p parts but there is one ring of p. If the current is needed to be controlled across the n channel, then the gate should be there around it. It is like putting a ring to the middle of a water flowing tube around. JFET is a device with three ends. They are called as source, drain and gate. The source and the drain are connected to the n type semiconductor whereas gate is connected to the p type semiconductor by Ohmic contacts.

First, let us take the voltage of the gate and the source V_{GS} as zero. Think that the source is earthed. As the drain is positive relative to the source, the electrons in the n type semiconductor piece flows from the source to the drain. Then how much is the potential of a point in the n type piece? There will be an Ohmic voltage drop across the n type semiconductor. The potential is linearly increased from zero (from the source) to the drain which is at a positive potential. Let us consider this as V_{DS}.

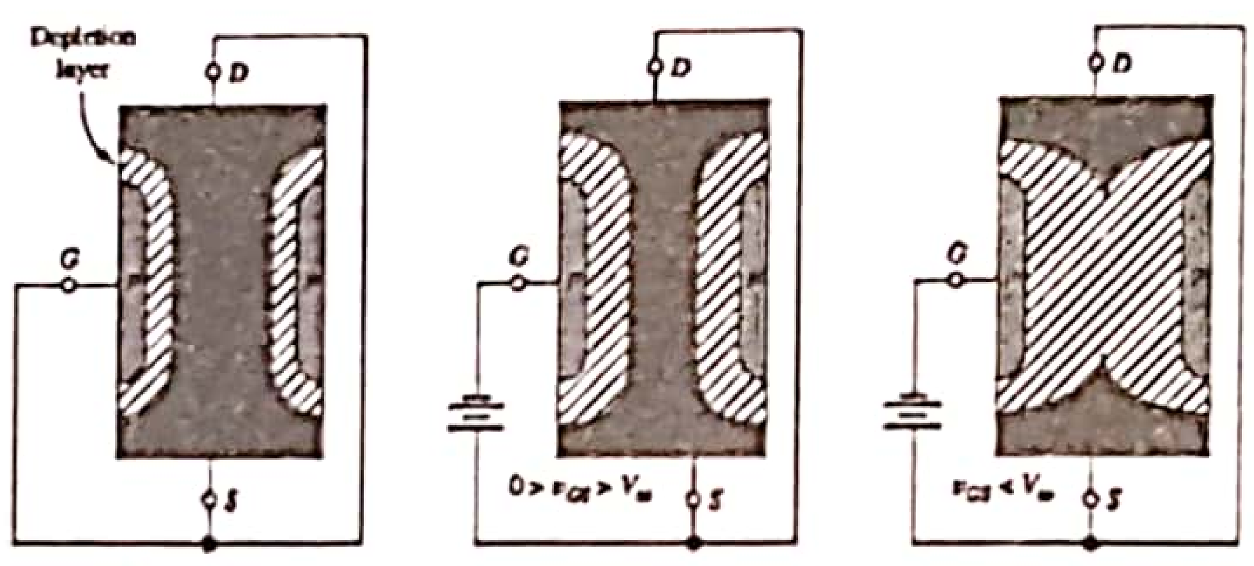
Now let us put our attention on the place of pn junction. First, we will consider $V_{GS} = 0$ instance. That means as the source is earthed, the gate is also earthed. So, the potential of p piece is zero. The n piece which is near to the p piece is positive. Then the pn junction is at reverse biased state. Hope you have learnt this under the function of a pn junction. Look at the figure.



In JFET, the pn junction (gate-source junction) is at reverse biased and this reverse biased voltage is gradually increased when it is going from the source to the drain. Why? As mentioned before, the voltage across the n type semiconductor is gradually increased from the source to the drain side. Due to this the depletion region created in the pn junction is bit wider when it goes to the drain side and when the gate is over, it should go back to zero state. The asymmetric shape of the depletion region has been shown in this figure. The drain current, I_D should flow across this depletion region. Even in two dimensions, you can see as two depletion regions. As p material is around the n material, there is one depletion region around. Think in a three-dimensional way.



Now, compared to the source, if you make the gate voltage negative ($V_{GS} < 0$), then with that change, the reverse biased voltage and the depletion region also get increased. When the depletion region is widened, the current (I_D) should flow across a thinner region (channel). When the current flow region is narrow, then the channel resistance gets increased. Therefore, when V_{GS} is made more negative, I_D is gradually reduced. So, at a certain V_{GS} value, the current flow can be made zero. This variation is shown in figure (1)..



If you think in another way, when the gate is increased negatively relative to the source, then the depletion region is widened and the drain current gets zero when it gets together at the widened place. It is like holding from your neck. It is like stopping the water flow in a rubber tube by pressing the fingers little by little. Figure (1) shows the characteristic V_{GS}-I_D in a n channel junction field effect transistor.

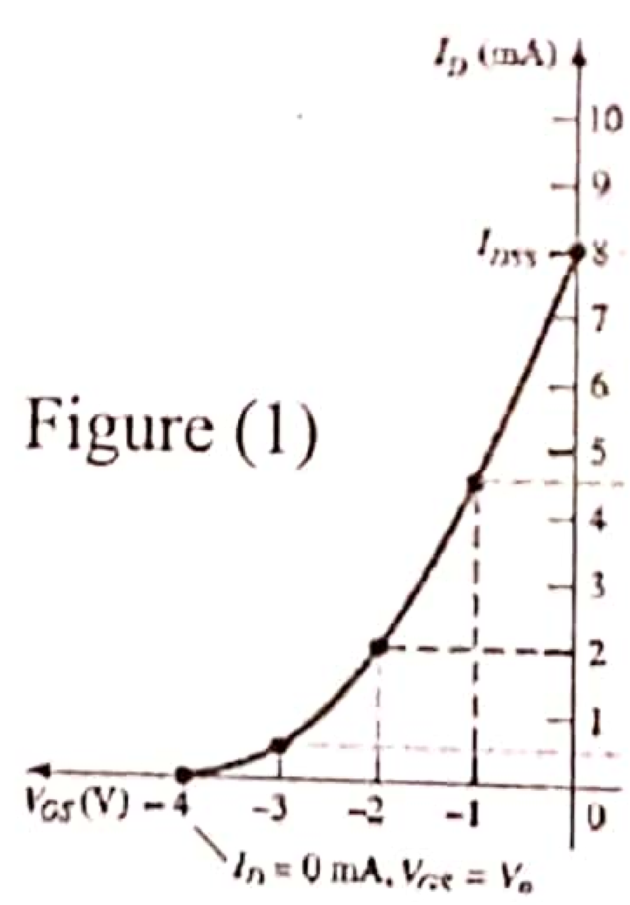


Figure (1)

Now, let us consider the characteristic of $V_{DS} - I_D$. This cannot be obtained directly. In a constant gate voltage (that means when V_{GS} is at a certain value), two effects are activated simultaneously.

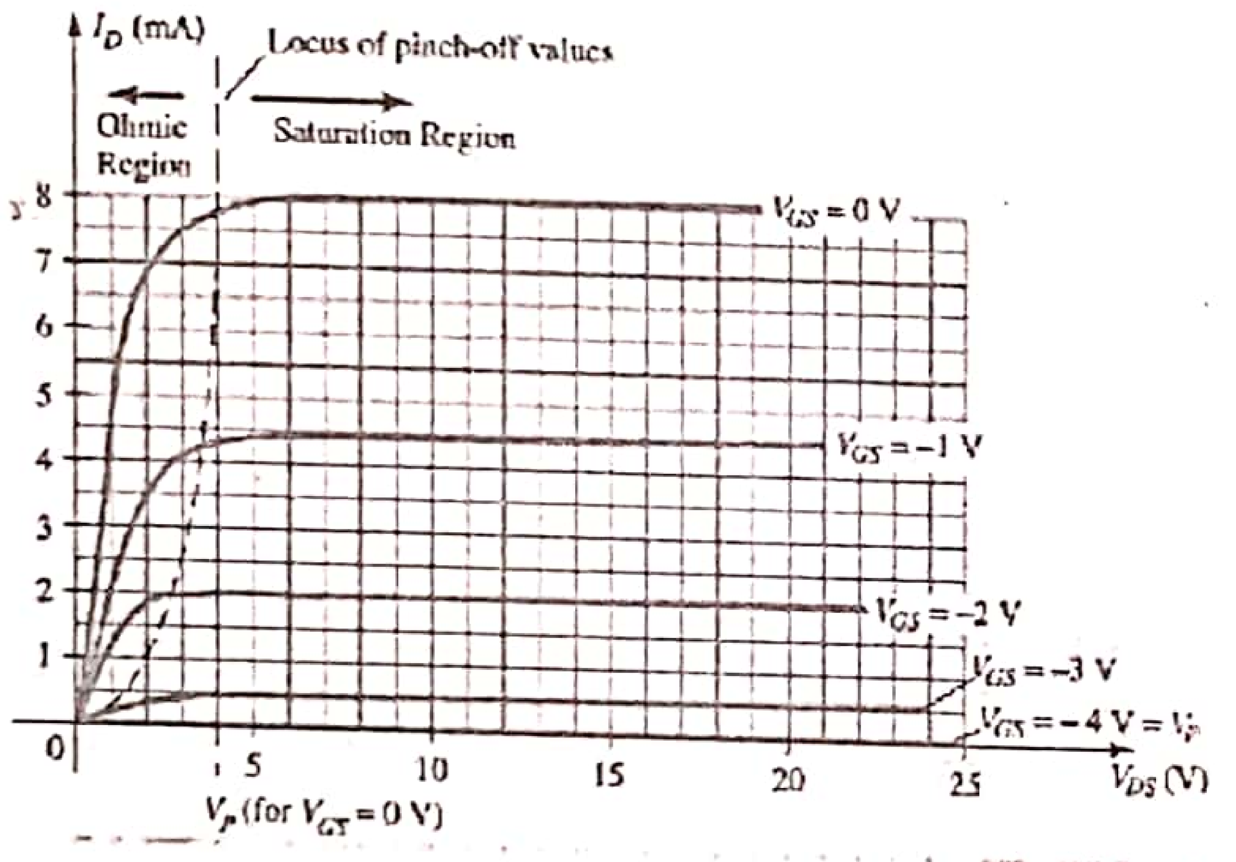
(i) When V_{DS} is gradually increased (when the drain voltage is increased relative to the source), then according to Ohm law I_D is increased.

(ii) But when V_{DS} is increased gradually, the reverse biased voltage across p-n junction is increased and then the depletion region gets widened. As the current is flown across the provided narrow channel, I_D is reduced.

These two effects are acted opposite to each other. So, will I_D increase with V_{DS}? Or will it reduce? You need to argue like this. When the depletion region is reduced in width and the channel is increased in width, if we consider that reduction of the width of the channel relatively small by increasing V_{DS}, then as the second effect is less active relative to the first effect, one can argue that initially I_D should be increased when V_{DS} is increased. But when V_{DS} is increased, then the width of the channel is greatly reduced. So, the result of the second effect gradually increases. Due to this the rate of increment of I_D with V_{DS} is gradually reduced whereas at a certain V_{DS} value, the effect on I_D by both of the effects cancel off with each other. Why? As from one effect it increases whereas from the other it decreases. The voltage where these two effects are being cancelled off is called as the pinch off voltage (V_P). After this value, I_D gets constant. Even we can say that the current is saturated. This characteristic is shown in figure (2).

When we grow old, our memory is also like this. In younger days, the things we put inside are greater than the things we remove. But once we get old, the things we put inside will be equal to the things we remove and our memory gets saturated. But love should not be saturated.

The value of the saturated current is dependent upon the voltage of gate-source (V_{GS}) You can understand this very easily. If V_{GS} =0, then at the beginning, then there is a narrow depletion



region. That means the width of the channel is greater. Therefore, the impact of the second effect is less at the beginning. But when V_{GS} is increased negatively (that means numerically reduced), as a greater reverse biased voltage is created at the beginning, the width of the channel is less initially. Then the impact of the second effect greater than the first effect. Due to this, the current is saturated at a lesser value. When V_{GS} = 0, the corresponding I_D value of pinch off voltage (V_P) is called I_{DSS}.

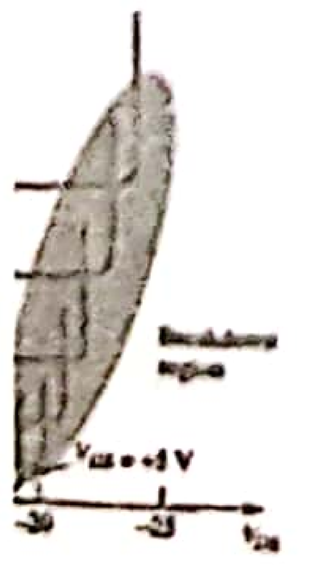
The above I_D- V_{DS} characteristic can be divided into four parts.

(i) Ohmic Region – When V_{DS} is increased initially I_D is increased. Then JFET is working as a normal resistor that is controlled by the voltage.

(ii) Cut -off Region – In this region, V_{GS} gets negative as needed and channel gets closed. JFET becomes to an open circuit state. That means I_D =0. This region is also known as pinch -off region.

(iii) Active or Saturation Region – In this region, I_D is being controlled by V_{GS}. For a constant V_{GS} value, I_D will not change with V_{DS}. When JFET is being used as an amplifier, it should be operated in this region. This saturated region should not be confused with the bi-polar transistor's saturated region. There when V_{CE} is nearly zero, I_C value gets saturated and the transistor gets into a saturated stage. The saturated region of a JFET is actually its active region. The word 'saturated' is being used because I_D is constant with V_{DS}.

(iv)

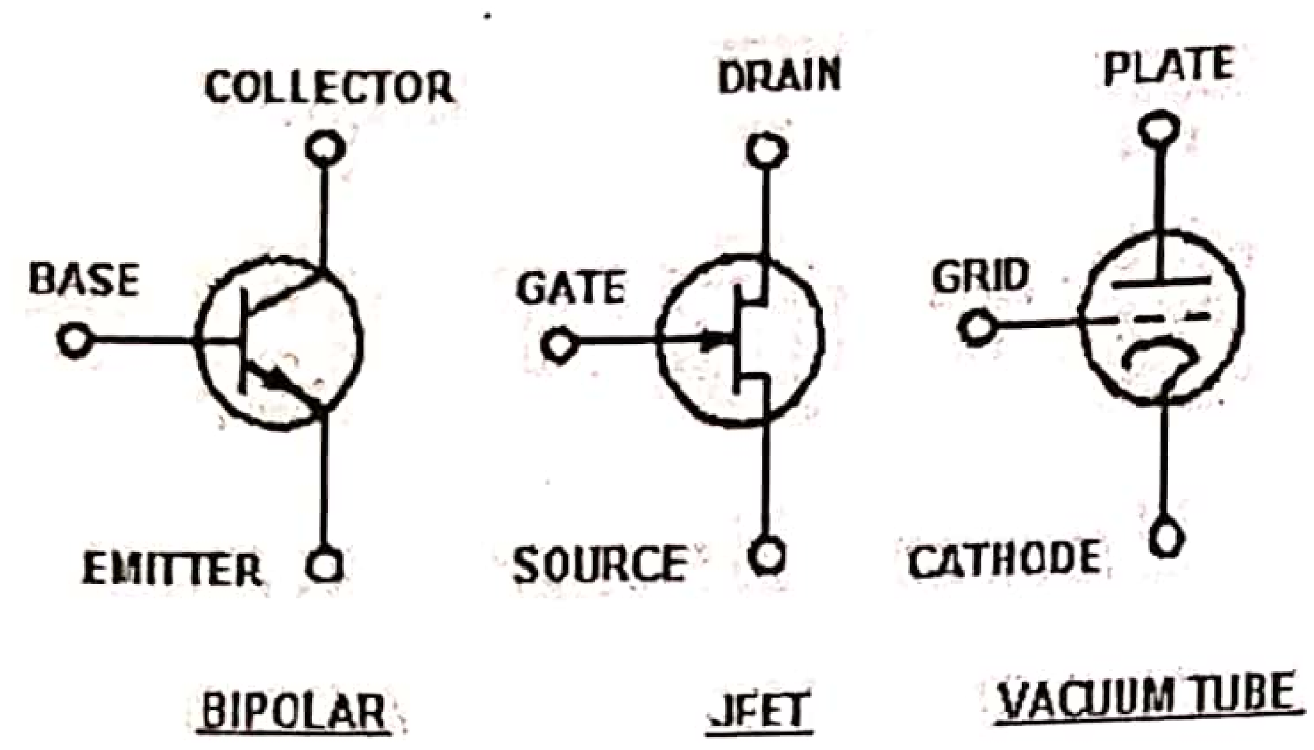


Breakdown Region – When V_{DS} is gradually increased, there will be an uncontrollable current flow across the channel and it will be broken down.

The dashed lines show the path of the pinch off voltages.

Using the symbols, given below is a npn transistor, n channel JFET and a vacuum tube. Type of equivalent ends in each of them is shown in the below table.

FET	Bi-Polar Transistor	Vacuum Tube
Gate (G)	Base (B)	Grid (grid)
Source (S)	Emitter (E)	Cathode (C)
Drain (D)	Collector (C)	Anode (A)



The vacuum tube has been mentioned for the comparison. You do not have to know about vacuum tubes. As I mentioned early, before the advent of semiconductor devices, what belonged to electronic were such vacuum tubes. Those days electronics were simple and easy compared to current days. As we found new friends, old friends should not be forgotten.

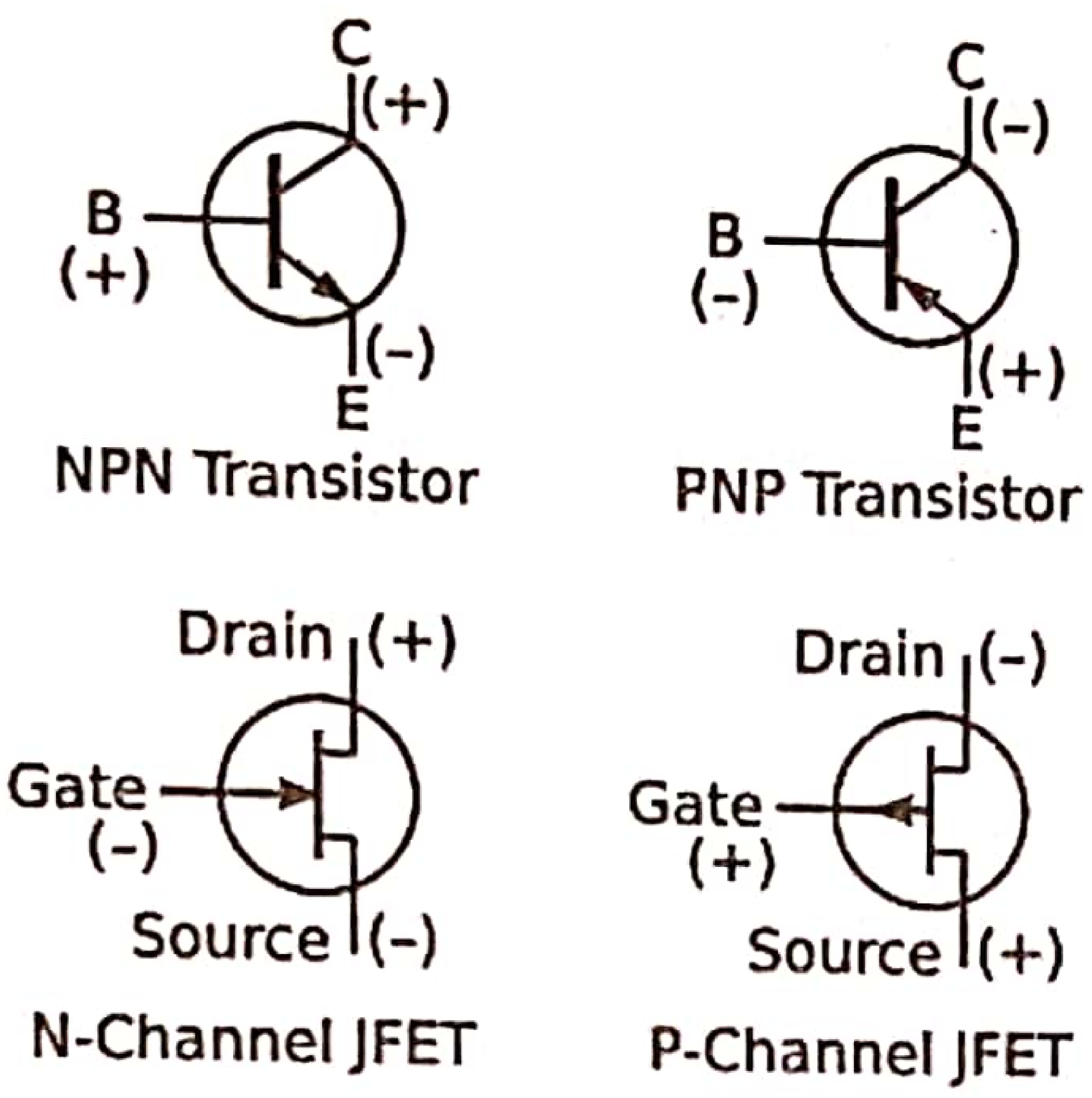
You need to realize that in a npn transistor the output characteristic at the common emitter configuration (V_{CE} against I_C) is mostly similar to the characteristic shown from the figure (2) of a JFET.

It is easy to obtain the characteristics shown in figure 1 and 2 from vacuum tubes. In a vacuum tube, electrons are emitted from the cathode. Electrons move from the cathode to the anode. The current flows from the anode to the cathode. To store electrons, the anode should be positive compared to the cathode. The grid is an electrode with holes. If the grid is kept at a zero potential, then maximum number of electrons will travel to the anode. By keeping the grid at a negative potential relative to the cathode, the electrons that travel to the anode can be controlled. If the grid potential is negative, then that means the electrons emitted from the cathode are being decelerated without being accelerated. It says do not come. Therefore, when the grid is kept more negative compared to the cathode, at a certain V_{GC} value the anode current I_A can be made zero. Is not this the characteristic of the figure 1?

Even the characteristic of figure (2) simply occurs in a vacuum tube. When $V_{GC} = 0$ and V_{AC} (the voltage across the anode and the cathode) is increased, initially I_A is increased. When V_{AC} is increased, I_A gets increased as the other electrons also come towards the anode. But when all of the electrons emitted from the cathode are collected by the anode, I_A will not increase as V_{AC} is increased. At a certain value of V_{AC}, I_A gets saturated. If all are being collected, then from whom are you going to take more?

Next, when the gird is made negative relative to the cathode, due to the deceleration given to the electrons (between C and G) the electrons that emit with lesser kinetic energy will not go towards the anode. When the gird is made more negative relative to the cathode, it forcefully says no to the people who come from the cathode. Those who come are decelerated and turned away. So, is it a surprise to see a gradual reduction of saturated current? The characteristic of V_{AC} and I_A is the variation of (2). Is not it? The characteristics of JFET correspond to the characteristics of vacuum tube. But the process that happens is completely different. Actually, the hope of Shockley was successful. He wanted to displace the vacuum tube from a semiconductor device.

The channel of p channel JFET is being made from a p type semiconductor and the gate is being made from a n type semiconductor also. <u>You do not have to know the function of a p channel JFET</u>. In the following figure, n channel and p channel JFET transistor symbols are being compared with npn and pnp transistors.



Always the arrow should be drawn towards the material of n type. It is the standard rule of drawing the arrow. In a bipolar transistor also, this rule is being applied. Look at the figure. According to this, there is one pn or np junction in a FET. Therefore, in n channel JFET, the arrow from p to n should be drawn on the gate inwards to it. In a p channel JFET, the arrow from p to n should be drawn on the gate outwards to it. There is no other alternative than drawing the arrow on the gate.

Now, let us look at the question. We need to find the statement/s that are not true. All the statements about npn transistor are true. If one is given as false, you do not have to look at JFET. Then the objective of the question is not being satisfied. Both (1) and (2) are correct. According to the long explanation above, n channel JFET keeps the gate-source junction in the reverse biased mode. Then only you can control the cross-sectional area of the channel. If it was forward biased, then there will not be depletion region.

Statement (3) is true for a transistor but according to a JFET that statement is wrong. The arrow is marked on the gate. (4) and (5) are correct. Only free electrons flow across n channel. In the functionality of a transistor, both electrons and holes participate. Therefore, transistor is a bipolar device. In the functionality of a FET only one from electron <u>or</u> holes participate. Therefore, FET is known as a unipolar device. In (5), I_B does not change depending on V_{BE} in a npn transistor. I_C changes depending on I_B. When looking at the variation shown above in figure (2), the expression given in (5) is valid for JFET.

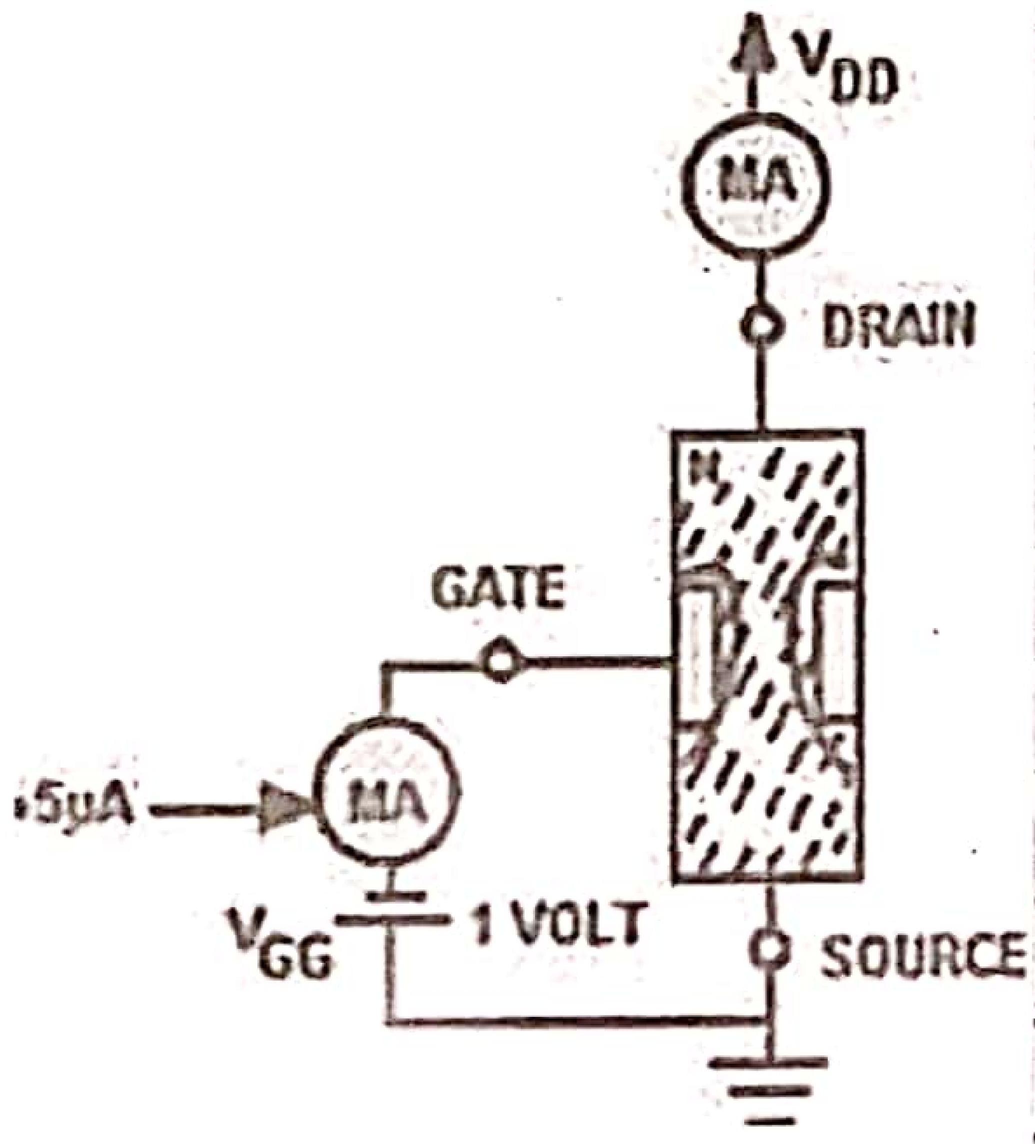
Finally, I hope that you are excited to know about the applications of FET. You know the applications of bi-polar transistors. Most of the time, FET is used as an electronic switch. You know that bi-polar transistor can be used as a switch using the cut-off and saturation regions. In such a bi-polar transistor, these instances are controlled by I_B current. (Look at the characteristic of I_B-I_C.) For a FET, if you look at the V_{GS}-I_D characteristic as shown in figure (1), then by controlling V_{GS}, you can see that I_D (channel current) can be changed from a maximum value to zero.

That means by applying a voltage, FET can be very easily used as a ON-OFF switch. Simply, FET is a device that is being controlled by a voltage like a vacuum tube. Therefore, FET is also known as a solid -state vacuum tube. Shockley used this term. A bi-polar transistor is being controlled by the current across the base and the emitter. When it is being used as a switch of ON-OFF, it is more effective to use a voltage for this purpose. By changing the voltage quickly switch process can be done from ON to OFF and OFF to ON. Therefore, FET is commonly used in switching circuits as well as logic gates. The other advantage is that in a FET, the energy consumption is less compared to a bi-polar transistor. So, the power creation of a FET is at a lower value.

Shown below are the typical values for a n-channel JFET. Consider that the source is being earthed. When $V_{GS} = 0$ and $V_{DS} = +5$ V, then $I_D = 10$ mA. The resistance of the channel (R) = 5/ $10 \times 10^{-3} = 500\ \Omega$.

When $V_{GS} = -1$ V and $V_{DS} = +5$ V, then $I_D = 5$ mA. Therefore, R = 1 kΩ.

From the above numbers we can see that, by a small gate-source voltage of 1 V, the resistance of the channel can be increased by double and the drain current has been cut down by a half. Another advantage of a FET is that its input resistance (between the gate-source junction) is at a higher value. Look at the following circuit.



When $V_{GS} = -1$ V and $V_{DS} = +5$ V, then the reading of the micrometer is 0.5 µA. From this you can see that the input resistance in between the gate-resource junction is $1/ 0.5 \times 10^{-6}$ =2 MΩ. What is meant as the input resistance is kept at a higher value is the current across input ends of the FET is very small when an input signal is given to the gate-source ends. That means the current drawn from the input is very little. This is a specific characteristic that should be possessed by the amplifier as well as many electric devices.

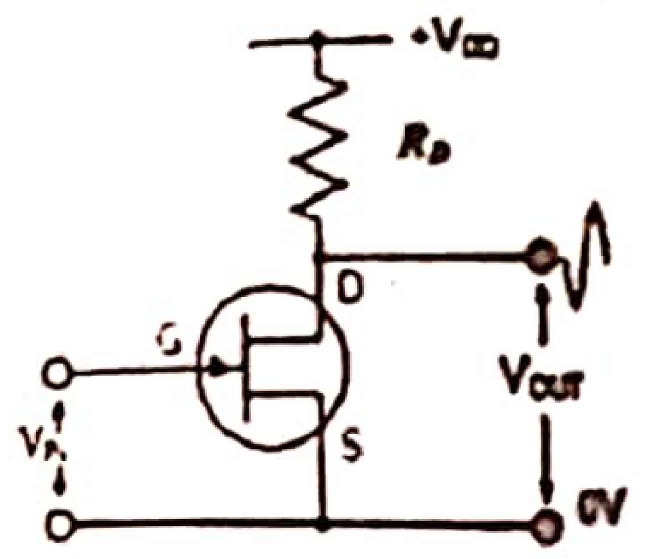
This characteristic is there is operational amplifiers. An ideal voltmeter has this characteristic. Here the nectar can be taken without crushing the flower.

In a normal bi-polar transistor, the input resistance between the base and the emitter is about 1 kΩ. Therefore, JFET has a higher input resistance. So, the resistance matching can be done easily when connecting couple of transistors with each other in a middle of a circuit. By connecting people with higher resistance there is no disadvantage to each other. One is not taking the things from the other unnecessarily. He is not waiting to eat from the other. As FET is small from the size, it needs a small space in integrated circuits (ICs). FET is also cheap in price.

Usage of JFET as an amplifier (common source configuration)

As there are three ends in JFET like in a bi-polar transistor, you can consider three configurations. They are common source (CS), common gate (CG) and common drain (CD). There is only common source - CS arrangement in the syllabus. This is shown in the figure. This arrangement is equivalent to the common emitter configuration of bi-polar transistors. According to the figure, the input is applied in between to the gate and source in this configuration. The output is obtained from the drain and the source. That means, $V_{IN} = V_{GS}$ and $V_{OUT} = V_{DS}$. The source is common to both input and output. That is why it is known as the common source configuration. As there is a greater input resistance between the gate and the source as well as a greater voltage gain can be obtained, this configuration is commonly used.

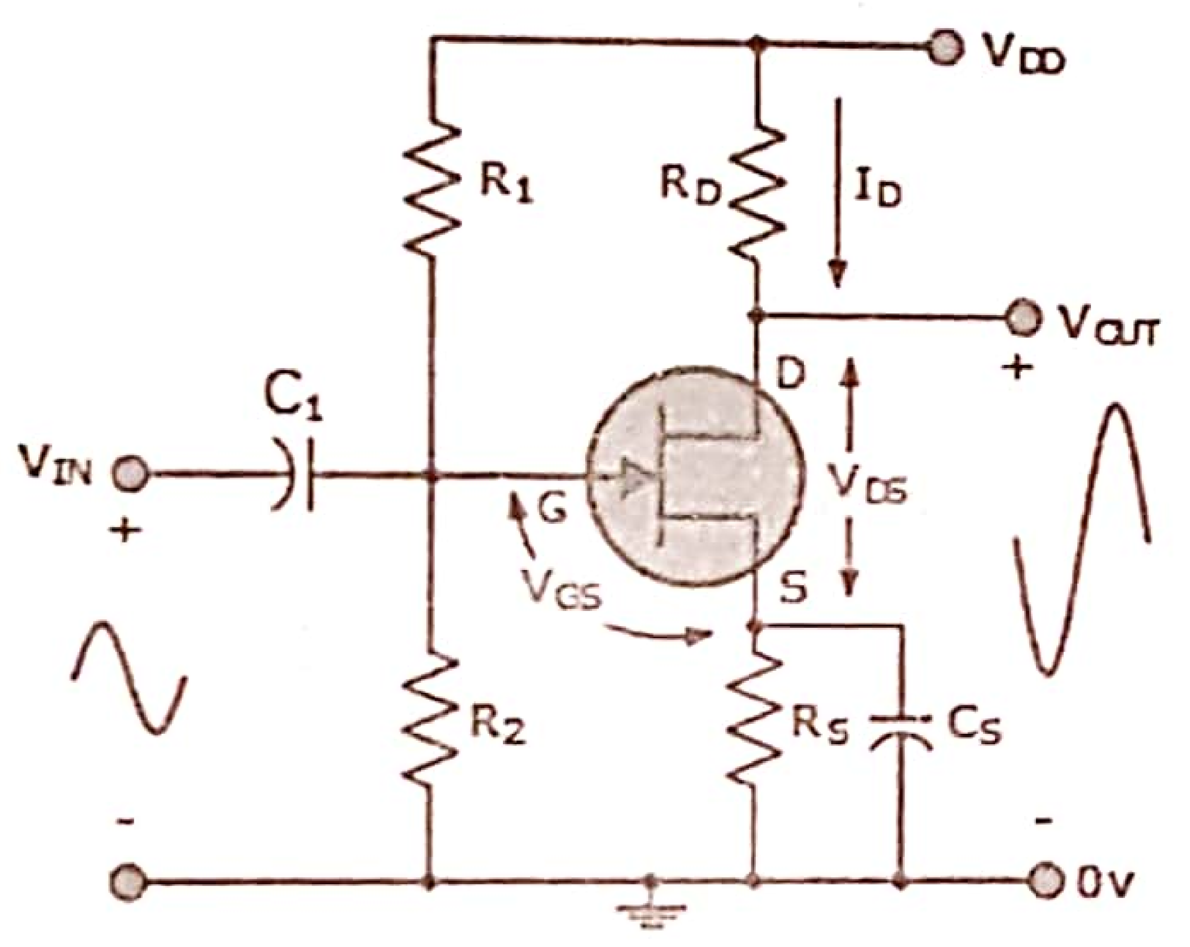
Even you can write equations for JFET amplifiers like bi-polar amplifiers.



$V_{OUT} = V_{DS} = V_D = V_{DD} - I_D R_D$

[can you remember $V_{CE} = V_C = V_{CC} - I_C R_C$?] According to figure (2) in a JFET I_D changes by a big amount for a small change of V_{GS} (here $V_G = V_{IN}$). According to the above equation, when relevant R_D values are selected, you can get the variation of V_{OUT} greater than the variation of V_{IN}. When I_D is increased V_{OUT} is decreased. When I_D is decreased V_{OUT} is increased. Therefore, like the common emitter amplifier, here also the output amplified signal has a phase change of 180^0 compared to the input signal.

As mentioned before, the specific advantage compared to the bi-polar transistor in common emitter configuration is the higher input resistance. When the input resistance is taking a higher value, then that amplifier is very sensitive to input signals. As the amplifier is not drawing a current from the input, any small change of the input can be felt by the amplifier. Therefore, this common source amplifier configuration of JFET is used commonly in audio frequency amplifier circuits. As long as we attract things of others, we too become insensitive people.



The figure shows a properly biased common source JFET amplifier circuit. This is equivalent to a common emitter amplifier circuit which is biased using the voltage divider method. The difference is that dc voltage gain $\beta = I_C/I_B$ in bi-polar transistor amplifier circuits is not interpreted in JFET amplifier circuits. As the input resistance of a JFET is at a higher value, it is always taken as $I_G = 0$. So, always for a JFET circuit it is $I_D = I_S$. In bi-polar circuits we take $I_C = I_E$ by considering I_B as small.

The other difference is that B-E junction of bi-polar amplifier circuits is in the forward biased mode. If it is a Silicon transistor, then we know that it is 0.7 V. In JFET amplifier, always GS junction is reverse biased. $V_{GS} < 0$ [$V_S > V_G$]. Also, that value should be given. We need to select a proper V_{GS} value to avoid distortions in the input and the output signals. It will be clear to you from the next graph shown. (Q point – quiescent point)

Other than the above mentioned two points, the rest of the calculations are similar to the bi-polar transistor amplifier circuit.

(1) As $I_G = 0$, $V_G = V_{DD} \cdot (R_2 / R_1 + R_2)$. From this you can find V_G.

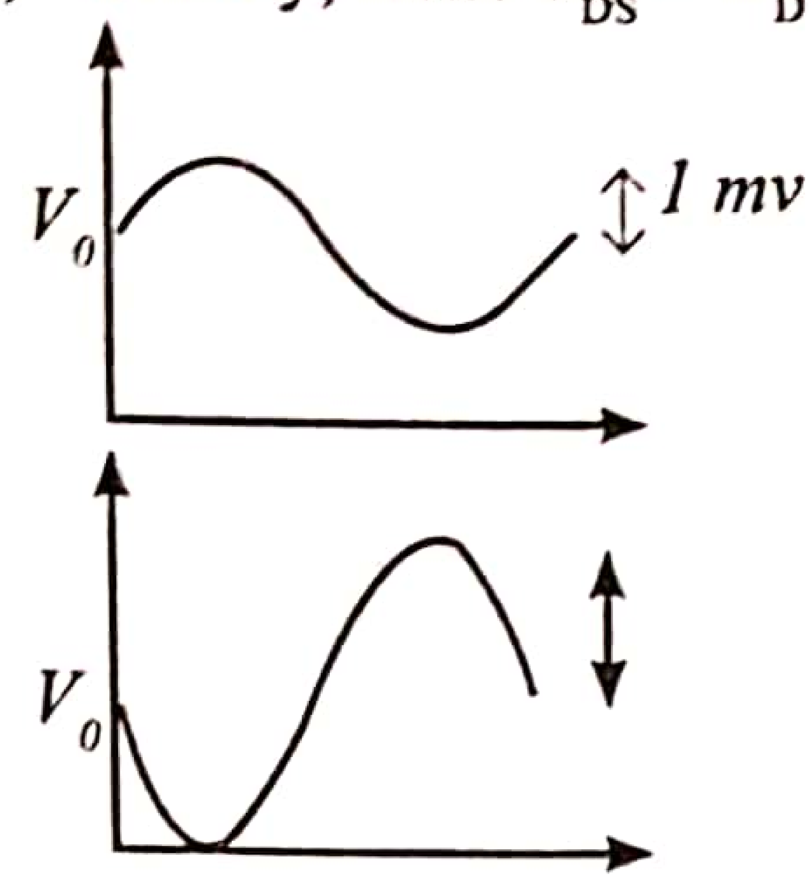
(2) Using $V_S = V_G - V_{GS}$ you can find V_S. Consider that $V_{GS} < 0$. Therefore, $V_S > V_G$. V_{GS} value should be given. It is not a standard value like V_{BE}.

(3) Now using $I_S = V_S / R_S$ and find I_S.

(4) $I_D = I_S$

(5) Next, use $V_D = V_{DD} - I_D R_D$ and find V_D.

(6) Finally, from $V_{DS} = V_D - V_S$ you can find the voltage drop across JFET.

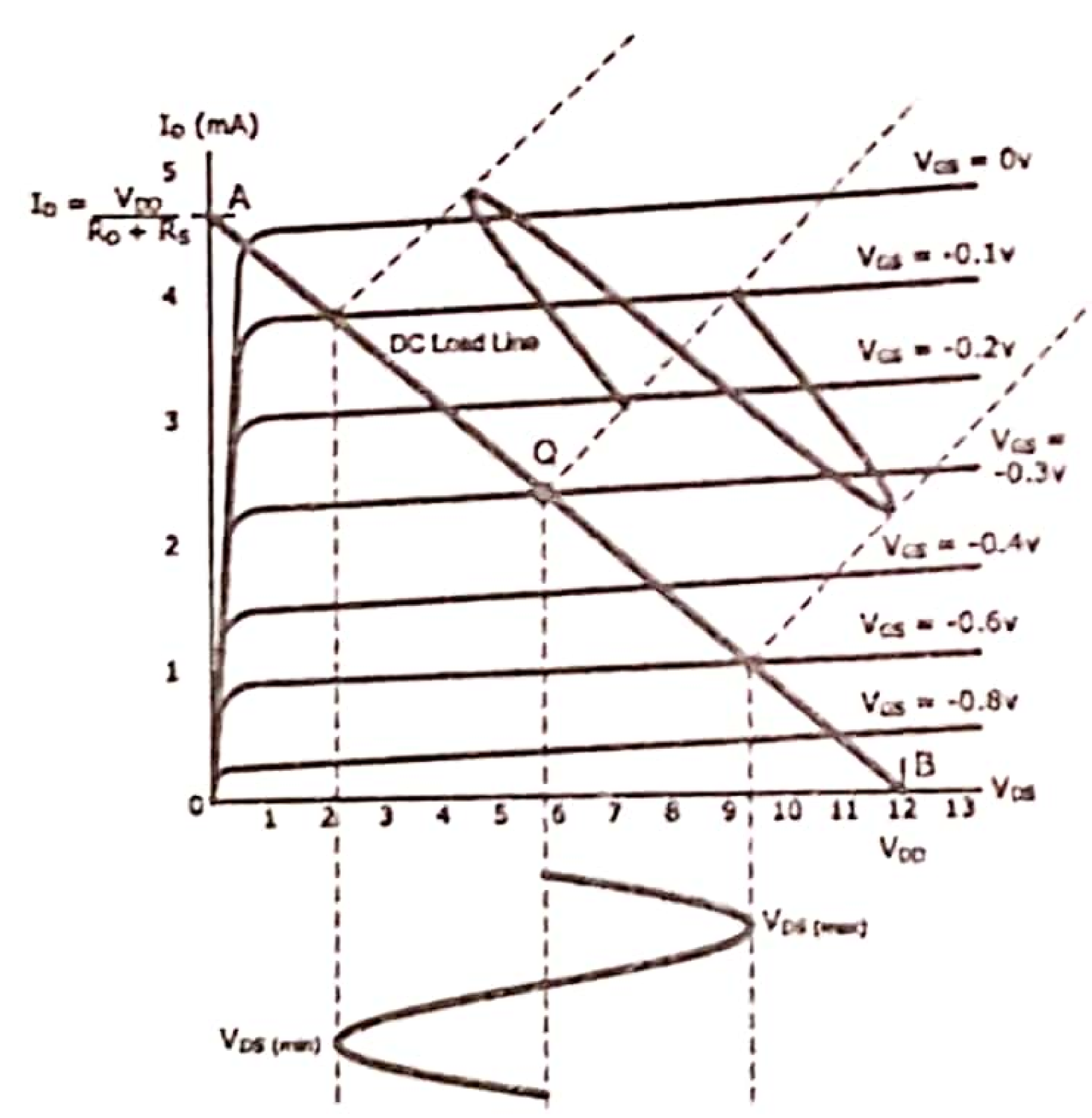


Now for example, think that an input signal voltage of 1 mV peak value has been applied to the gate end. Then across C_1 input capacitor there shows a superimposed 1 mV sinusoidal signal on V_G dc voltage at G end.

If A is the alternating voltage gain of the amplifier, then what is created on D end is an alternative output signal with a A X 1 mV peak voltage which is superimposed on dc value of V_D. The expressions for alternating voltage gain for bi-polar transistor amplifiers and JFET amplifiers are not in the syllabus. Therefore, in a problem, A should be given.

The blocking capacitors act is same. The biased state of the dc currents is maintained by them. As dc currents do not flow across the capacitors, they block the part of the current that flows across R_1 and R_2 which can flow into the input signal source and they block the part of the current of I_D which can flow towards the output signal. As shown from the figure is the unmutated input and amplified output signals with the load line. The load line can be obtained by $V_{DD} = I_D (R_D + R_S) + V_{DS}$. When I_D is taken, $I_D = -\left(\frac{1}{R_D + R_S}\right) V_{DS} + \frac{V_{DD}}{(R_D + R_S)}$

The variation of V_{DS} against I_D is a straight line. The gradient is $-\left(\frac{1}{R_D + R_S}\right)$ whereas the intercept is When $V_{DS} = 0$, then $I_D = \frac{V_{DD}}{(R_D + R_S)}$. ', then $I_D = V_{DS} = V_{DD}$. When $I_D = 0$, then $V_{DS} = V_{DD}$.

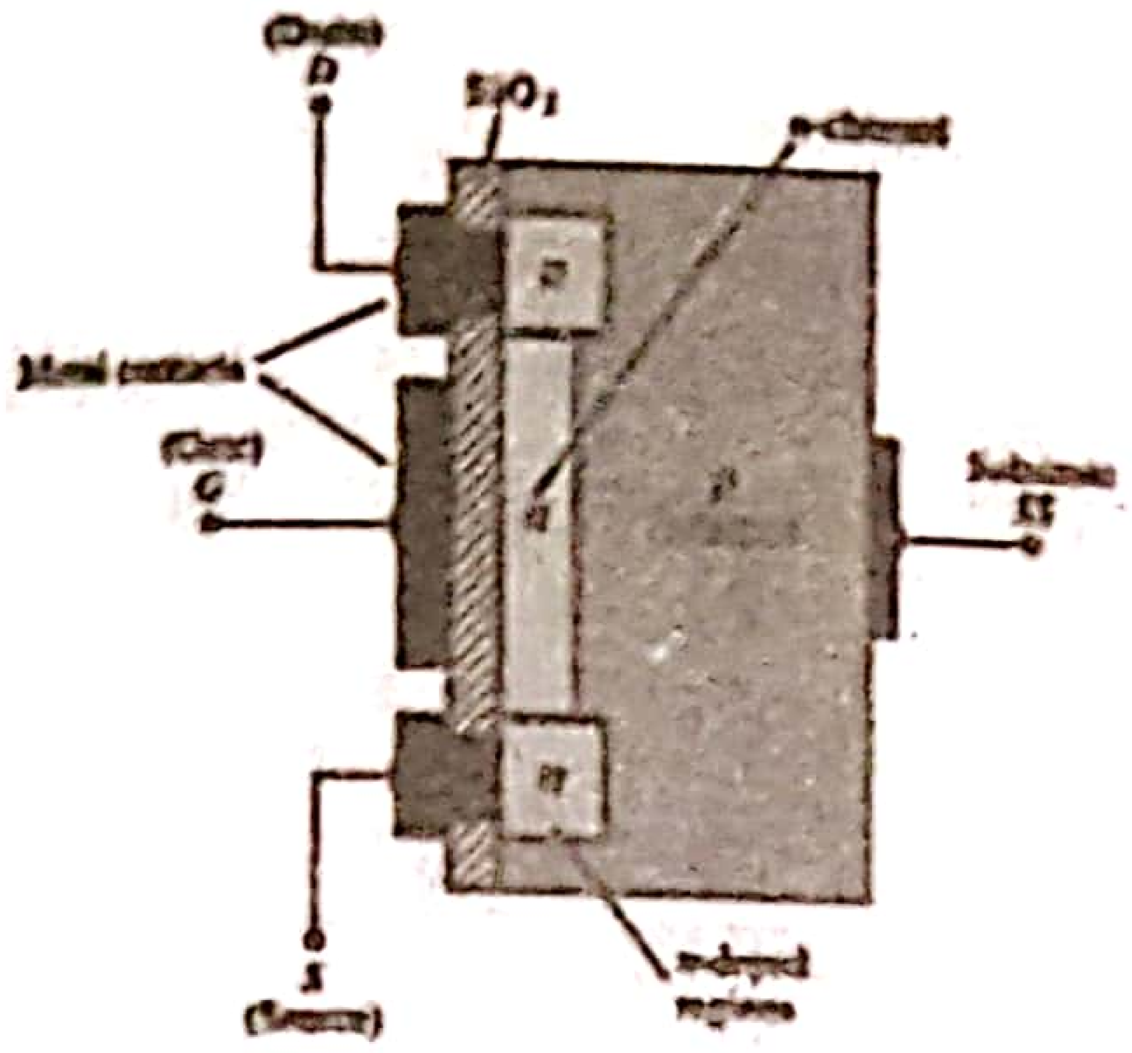


As in a common emitter amplifier, at quiescent point (Q point) $V_{DS} = ½ V_{DD}$. That means V_{GS} should be chosen to the middle of Q point. For the shown instance, $V_{GS} = -0.3$ V. The resistor R_S value helps to keep V_{GS} at the correct value. Earthing of the alternative current part of I_S current due to the input signal is done by the capacitor CS. The alternative current part goes across C_S without going across R_S. Therefore, the voltage across R_S does not change according to the signal.

MOSFET (Metal Oxide Semiconductor Field Effect Transistor)

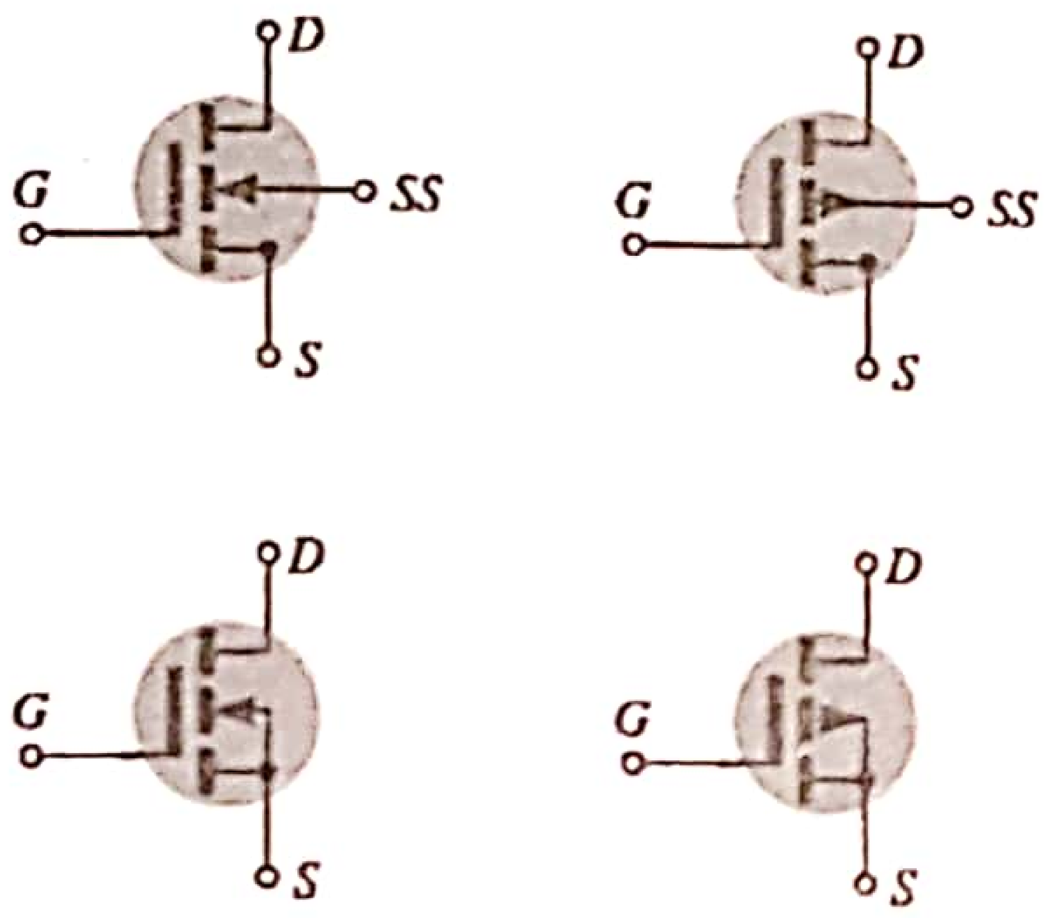
To discuss the main differences of MOSFET and JFET is only there in the syllabus.

MOSFET is functional in two modes. In both ways, a thin oxide layer is deposited (most of the time SiO_2 is used) in between the gate and main channel that carry the current. Due to this thin plate, oxide term has been used. The term metal is there as the ends are keeping connections through metals (normally Al).

(1) 

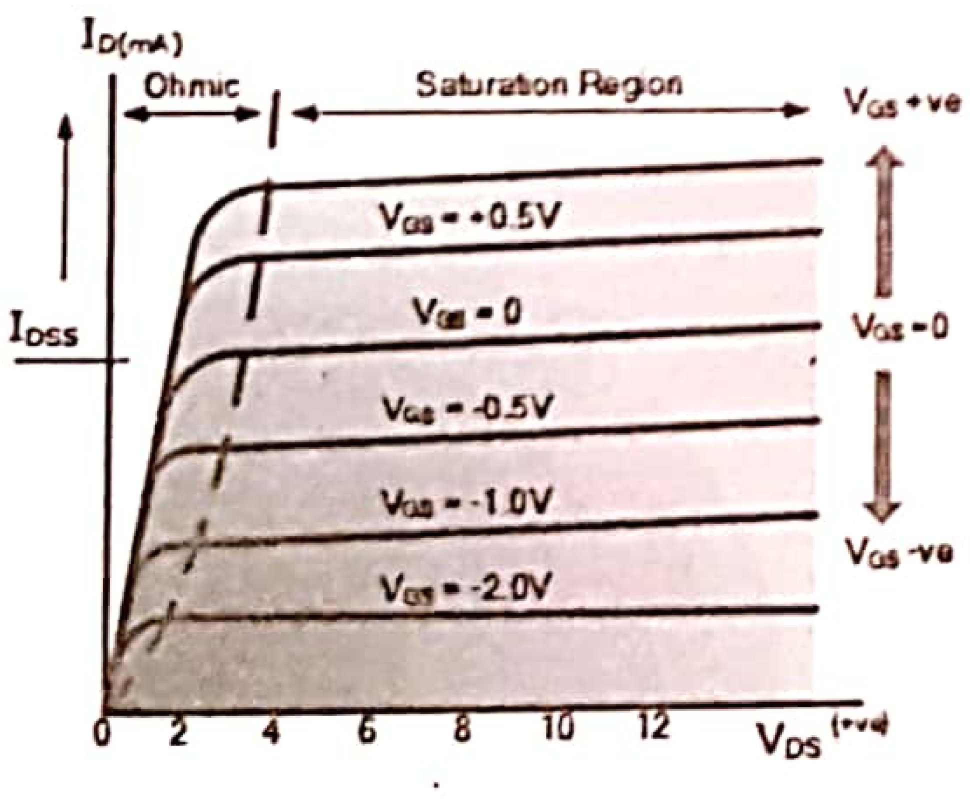
The configuration of n channel MOSFET working in the depleted mode is shown in the figure. This is equivalent to n channel JFET in every way. Drains and source leads are being connected to two regions doped with n. These n doped two parts are being connected across a n channel. This n channel is connected to the gate across a thin SiO_2 insulating film (10^{-4} mm). Regions doped as n are on a substrate which is doped as p. In this subtrate which is doped as p has an end (SS) that can be used to get an extra connection. This end is almost connected to the source internally. When SS end is earthed, then the source automatically gets earthed.

A symbol with a depeltion type MOSFET is shown in the figure. By showingeither SS end or by connecting SS end to S the relevant symbol can be represented.

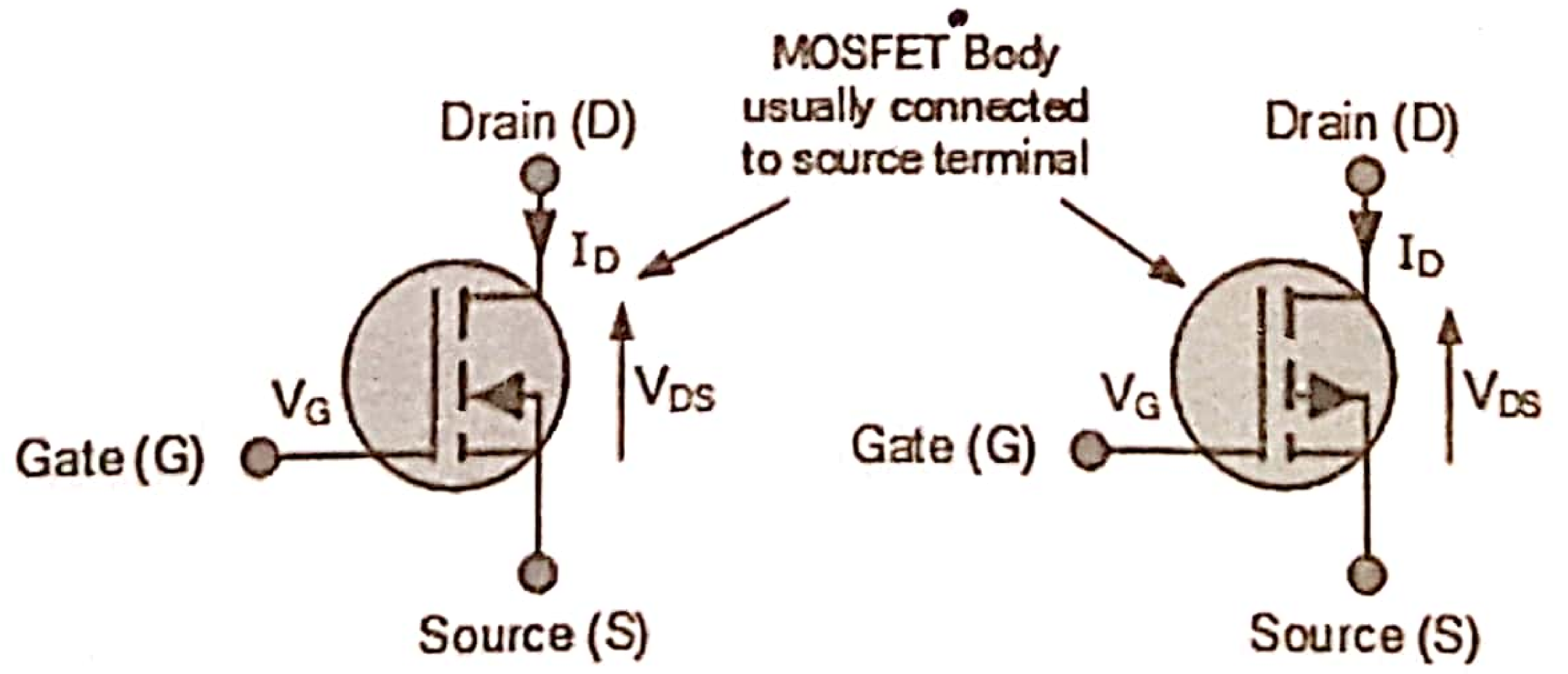


I_D- V_{DS} characterisitic of such a depletion type n channel MOSFET is shown in the figure. It can be seen that it is almost similar to the corresponding characteristic of a JFET. The main difference is that V_{GS} (>0) is not kept at a positive value at any instance in JFET. If V_{GS} gets positive, then p-n juction will be at the forward biased mode. Then the electrons will not go towards the drain but will go towards the gate. Therefore, for the functionality of a JFET, always VGS should be kept at a negative value.

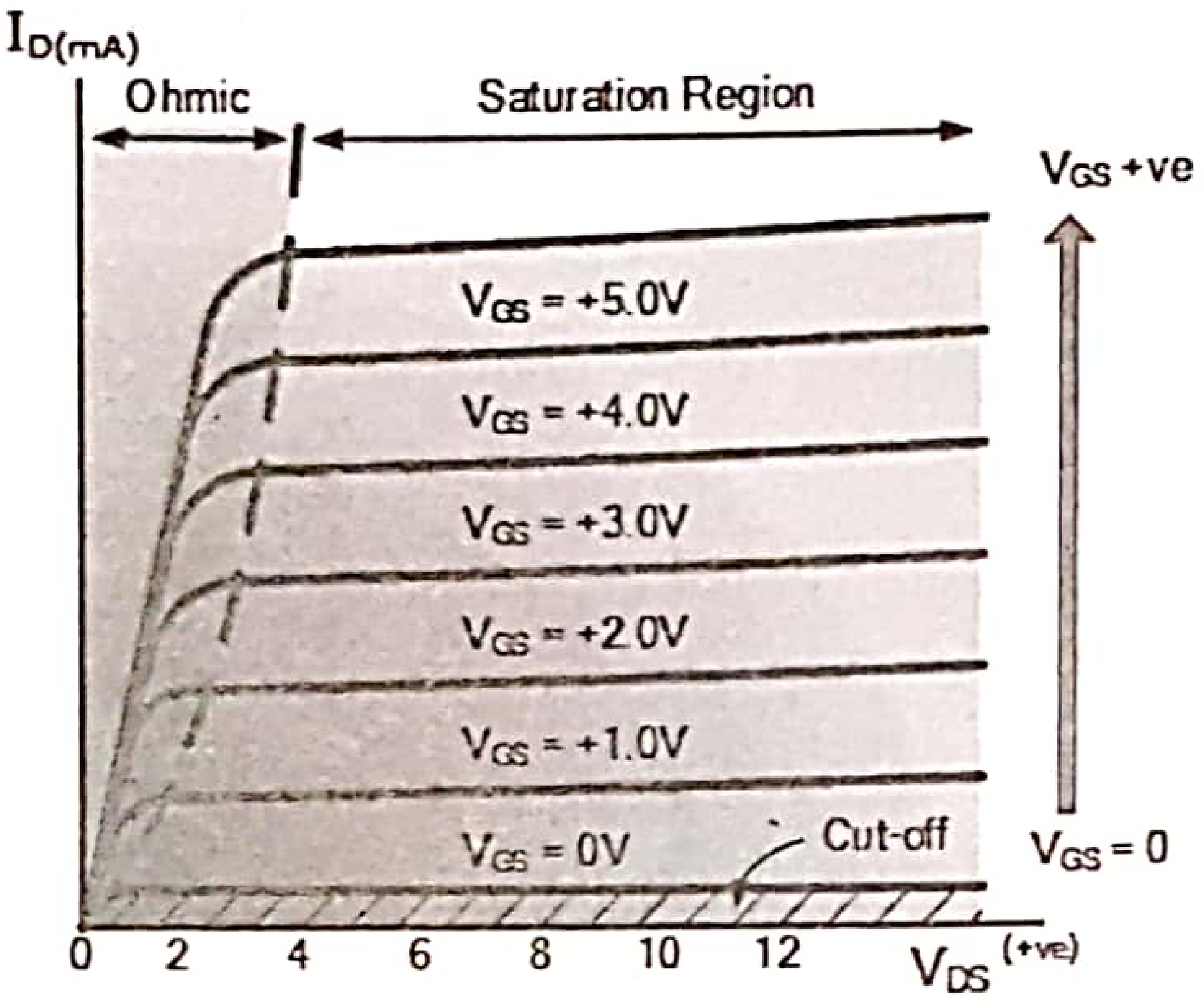
But if you need, then V_{GS} cannbe made positve in a n MOSFET. As the oxide layer between the channel and the gate acts as an insulator, the produced electrons are not reached to the gate. They are being turned back to the side of the drain. Therefore, when V_{GS} is made positive, you cannot expect a development in I_D. When all the electrons are being collected, I_D gets saturated.

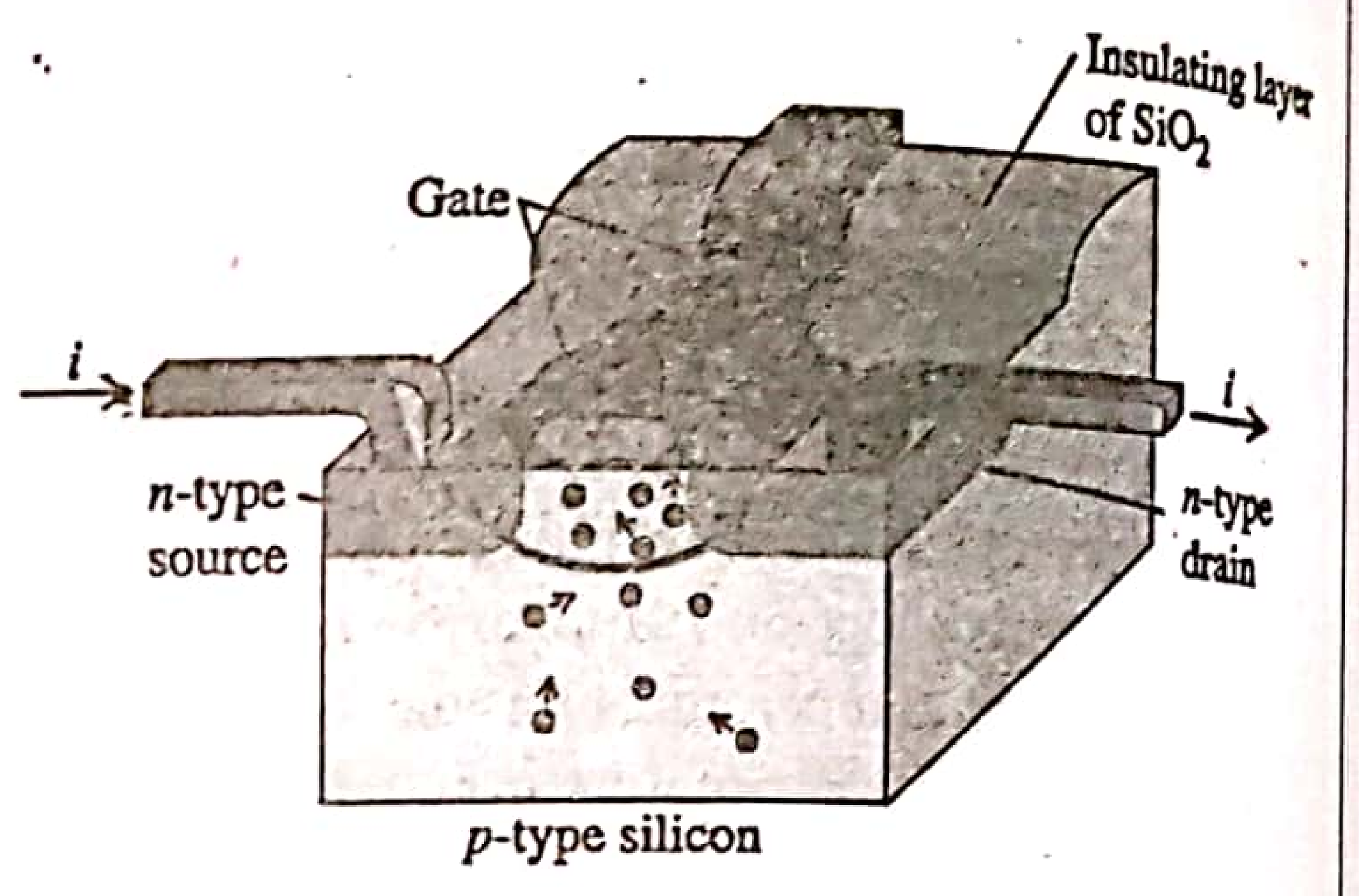
(2) 

Mainly, MOSFET is used in enhancement mode. The configuration is different than the earilier case. The figure shows such n channel configuration. Like before, the drain and the source are put into the region where it is being doped as n. But between these two regions there is no n channel as before. What is there in between these two is a p type substrate which is doped a bit as p. The other connections are as before. The symbols are shown here.



I_D -V_{DS} characteristic has been shown of a such n channel MOSFET. We will try to understand this. In this mode, every time V_{GS}(>0) should be positive always. The argument is like this. In the p type material, there are no such free electrons but there is a little. When V_{GS} passes a threshold value (V_T), then the electrons in the p substrate will be attracted to the gate. Look at the figure. If you think in another way, the holes in the p type subtrate are being repulsive by the gate. Due to this reason, the electron concentration near the gate gets increased. But due to the oxide layer, they are not travelled to the gate. These electrons create an electron channel from the source to the drain. These electrons create a bridge between n type source and n type drain.





But when all the electrons are being collected, at a certain V_{GS} value (>0), I_D gets saturated like before. When V_{GS} is increased more positively, the saturated current is increased as more and more elctrons are snatched from the p substrate. The substrate is being less doped to snatch many possible electrons.

Here you will understand that there is no usage by making V_{GS} negative (<0). When V_{GS} is made negative, then the remaining electrons will go away from the gate. Then there will not be a bridge between the two islands of n type.

The main advantage of a MOSFET like this is I_D gets zero when V_{GS} =0. The main advantage of a JFET or depletion type MOSFET is that there is a current I_D even when V_{GS} =0. When a MOSFET working in the increment mode gets V_{GS}=0 (that means without any biased mode) as I_D gets zero, then it automatically work as an open switch (OFF). When V_{GS}> V_T then it works as a closed switch (ON). Therefore, in many electronic circuits with switches, the memories of logic gate circuits, calculators and computers, MOSFET working in the increment mode is being used. The V_T value of n cahnnel devices are in the range from 0. 5 V to 0.8 V.

Therefore, very efficiently it can exchange between 'ON' (closed) and 'OFF' (open). When V_{GS}> V_T (it will be 'ON' whereas when V_{GS}=0, it will be 'OFF'.

In a JFET or a depletion mode MOSFET, when V_{GS} =0 it will be 'ON'. If you need to 'OFF' then V_{GS} should be -4 V or -5 V.

A MOSFET working in the increment mode can be also used as an amplifier in the common source configuration. Due to insulating oxide layer its input resistance is higher than a JFET ($10^{12}\Omega$ - $10^{15}\Omega$). Therefore, it is far better. The power consumption of these are taking a very low value too. There fore, these are being used in microprocessors.

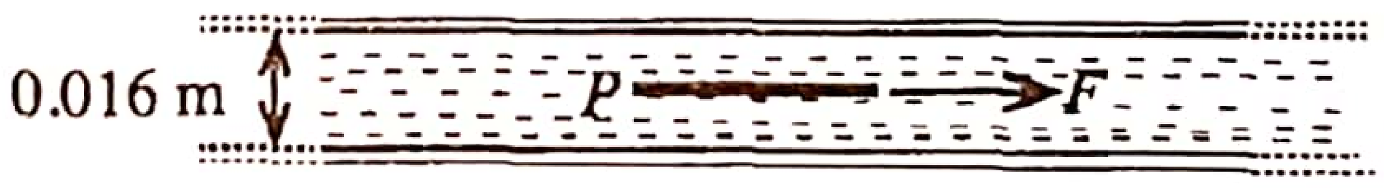
The channel of a MOSFET in depletion mode is shown in an unbroken line as I_D is not zero even when $V_{GS} = 0$. In a MOSFET working in the increment mode, the channel has been shown in a dashed line because $I_D = 0$ when $V_{GS} = 0$. There is a space in between the gate and the channel to show the insulating oxide layer. Check the rule for the arrow in the symbols are correct. Always the arrow is being drawn from p to n. Here the arrow cannot be drawn on the gate. Type p is at the substrate.

I think it is better to put a note on Shockley and end this long description. Due to the new discoveries of Shockley and his companians the foundation was laid for the semiconductor devices manufacturing industry which is famous as 'Silicon Valley' in California, USA. Shockley had contributed Silicon to become a "Silicon Valley'. But due to very strict views, his popularity was decreased during his latter periods of his life. He expressed that the future generations of human species should be made only from the genes with skills. He said in public that people with genes of less skills should not reproduce many off-springs to the world.

There is a rumour that the sperms of Nobel Laureates are deposited safely in a gene bank. Shockley had mentioned in public that he had given his sperms for this bank. 99% of Nobel Laureates for Physics are male either for good or bad!! It may be that males think and do unwanted things!! But only males cannot reproduce children alone. 216 out of the Nobel Laureates for Physics (upto now by 2020), there are four females. One of them is the very famous Marie Curie.

When Shockley was dead, it was said that his death was known to his children from the newspapers. This is the nature of the western culture. The world really needs genes which identifies love and affection even there are no special skills. As I think, it is enough to have less number of people with special skills. When there are more people with special skills, they tend to have conflicts. It is better if there can be more people with characteristics of kindness and caring nature.

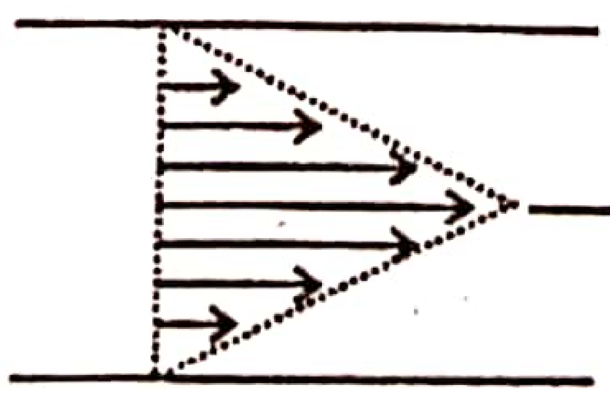
35. **Figure shows a section of a long horizontal rectangular tube of height 0.016 m having a large surface area, and filled with a lubricating oil of viscosity 0.072 Pas. What is the force F required to drag a very thin plate P of area 0.4m² with a velocity of 0.02 m s:' along the middle plane between the top and bottom surfaces of the tube as shown in figure?**



(1) $3.5\pi \times 10^{-3}$ N (2) $7.0\pi \times 10^{-3}$ N (3) 3.6×10^{-2} N (4) 7.2×10^{-2} N (5) 1.44×10^{-1} N

Viscosity 10

It is very simple. The only difference is that the plate is being carried in the middle of the tube. Normally, the plate is being carried on one liquid level. There are two layers of oil on the top and the bottom of the plate. We can consider that oil layers (the top and the bottom) that touch the plate is going in the same speed of the plate and speed of the oil layer that touches the internal surface of the tube can be considered as zero. The velocity profile of the thin oil layers in the tube are as follows.



As the velocity is uniform, the applied force is equal to the total viscous force. When finding the viscous force, you need to consider both sides of the plate.

Use the standard equation for the viscous force which is $\eta A \Delta V/\Delta r$.

Needed force = 2 X 0.072 X 0.4 X (0.02/0.008) = (2 X 72 X 0.001)/ 8 = 0.144 = 1.44×10^{-1} N.

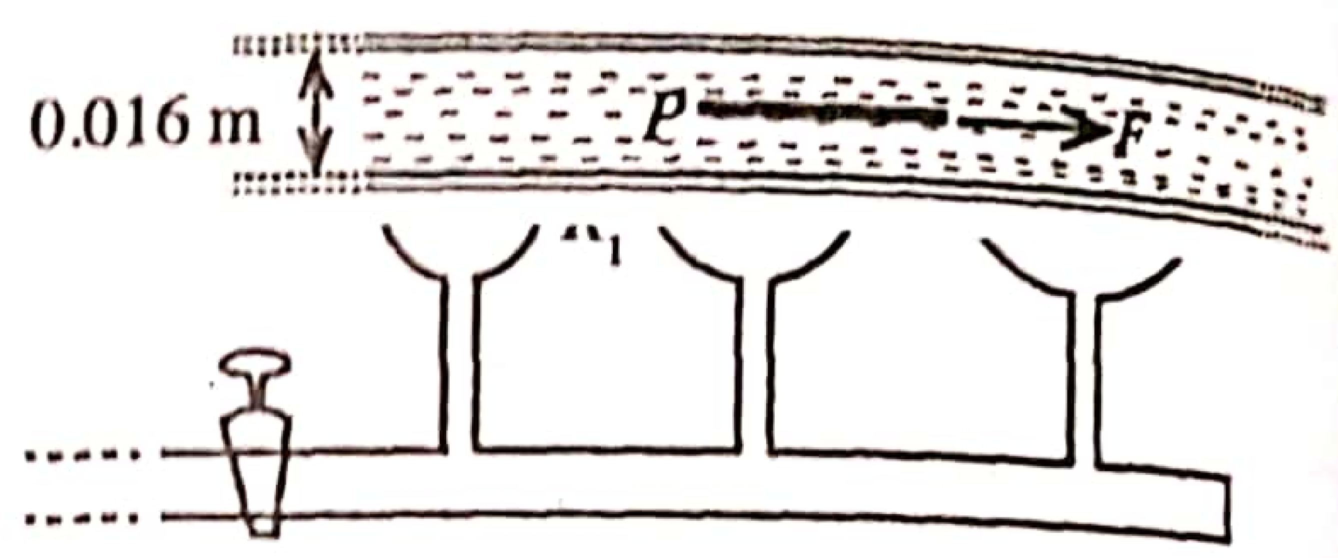
You can get it wrong if you do not consider the viscous force that is affecting on both surfaces of the plate. You need to consider that there is an oil layer under the plate as well as the top of the plate. To neglect the edge effects on both sides of the tube, it has been mentioned that the tube is with a big surface area.

36. Three spherical liquid films of surface tensions T_1, T_2 and T_3 respectively are in equilibrium as shown in figure such that the corresponding radii $R_1 = r$, $R_2 = 2r$ and $R_3 = 3r$. Then,

(1) $T_1 = T_2 = T_3$ (2) $\frac{T_1}{3} = \frac{T_2}{2} = T_3$

(3) $\frac{T_1}{6} = \frac{T_2}{4} = T_3$ (4) $T_1 = \frac{T_2}{2} = \frac{T_3}{4}$

(5) $T_1 = \frac{T_2}{2} = \frac{T_3}{3}$



Surface Tension

There is nothing to calculate. The pressure outside the spherical liquid film is same. Even the pressure inside the spheres is same. If so, $T_1 = T_2/2 = T_3/3$. What else should have happened? The extra pressure is 4T/R. As 4 is common, it does not need to be written. If the surface tensions are equal, then the spheres cannot stay like this way. The pressure is increased in the smaller sphere. If so, then the smaller spheres will shrink and the bigger sphere will get larger. But as the surface tensions of the liquid films are different, there is no issue for being in equilibrium for such a system.

I remember alveoli in the lungs. Big and small alveoli can be in equilibrium by connecting with each other. As there are more liquid surfactant molecules which reduces the surface tension in a unit area of small alveoli, its surface tension has been reduced. By that the disadvantage of increased pressure due to reduction of radius has been rectified by the reduction of surface tension. $T_1 < T_2 < T_3$.

37. A cylindrical copper block- of radius r and length $l = 2r$ radiates energy as a black body at temperature T. If this copper block is cut and separated into N identical disks having the same radius r, the rate of the emission of radiant energy at the above temperature will increase by a factor of

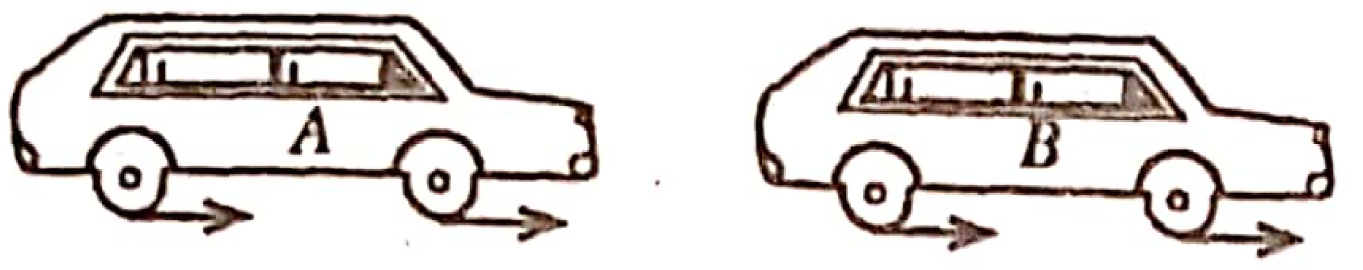
(1) $\frac{(N+3)}{3}$ (2) $\frac{(N+2)}{3}$ (3) $\frac{(N+1)}{3}$ (4) $\frac{N}{3}$ (5) N

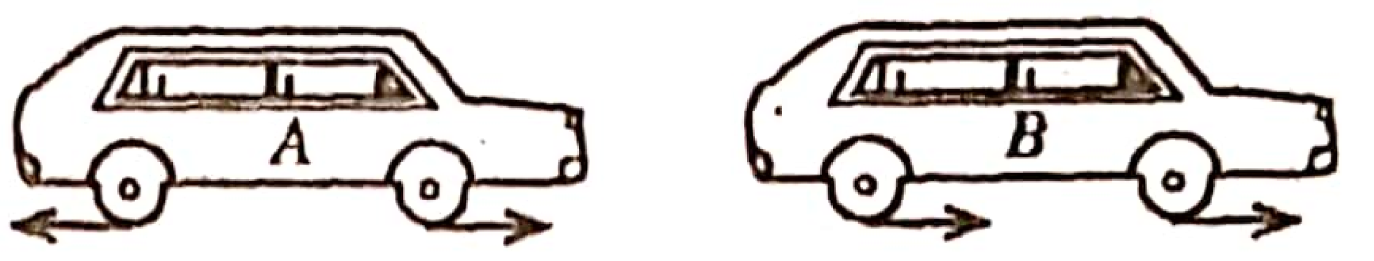
Radiation

As the temperature is not changed, the emitting energy is dependent upon the total surface area. The total surface area of the cylindrical block is $2\pi r^2 + (2\pi r \times 2r) = 6\pi r^2$. $2\pi r^2$ is the area of the two plane sides. $2\pi l$ is the area of the curved surface. The total area of the curved surface does not change as the block is cut into several disks. It is equal to the previous value. But when the disks are being cut, the net area of the exposed plane surface gets increased. The total area of the two plane surfaces in a disk is $2\pi r^2$. Therefore, if there are N disks, then the total area of the plane surfaces is $2\pi r^2 N$. The total area of the curved surface of N disks is equal to the previous area of $4\pi r^2 (2\pi l)$. Therefore, once the disks are cut, their total exposed area is $2\pi r^2 N + 4\pi r^2 = 2\pi r^2 (N + 2)$. $2\pi r^2 (N + 2)$ is (N+2)/3 times of the previous area of $6\pi r^2$.

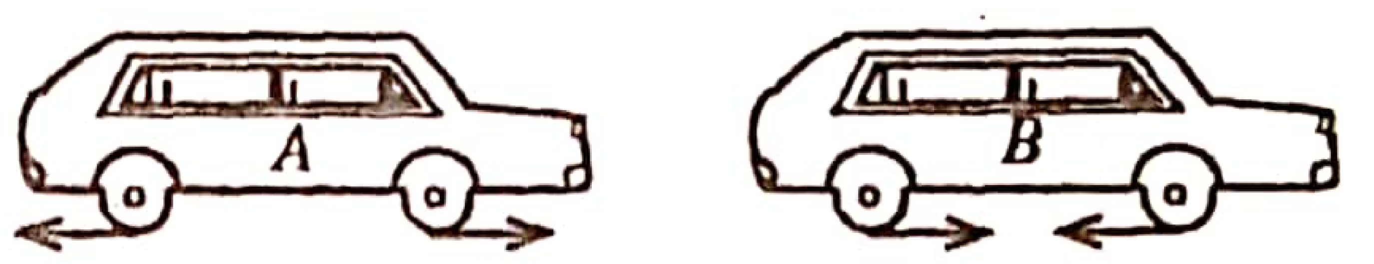
This cannot be solved without calculation. After the disks are cut, if you do not see that the total of curved surface area in the disks is equal to the previous curved surface area, then you will take much time to get the answer. This is the trick here. Even if you do not see, the curved surface areas of the disks can be also found like this way. As N disks with equal thickness are being cut, the thickness of a disk is $l/N = 2r/N$. Then the curved surface area of a disk is $2\pi r.\ 2r/N$. Therefore, the total curved surface area of N disks is $2\pi r.\ 2r\ N/N = 4\pi r^2$

38. Consider two motor vehicles, A and B. In motor vehicle A only the front wheels are coupled to the engine and rotated; and in vehicle B only the rear wheels are coupled to the engine and rotated. Which of the following diagrams correctly shows the directions of the frictional forces acting on the front and rear wheels of motor vehicles A and B by the ground, when they are travelling in the forward direction?

(1)


(2)


(3)

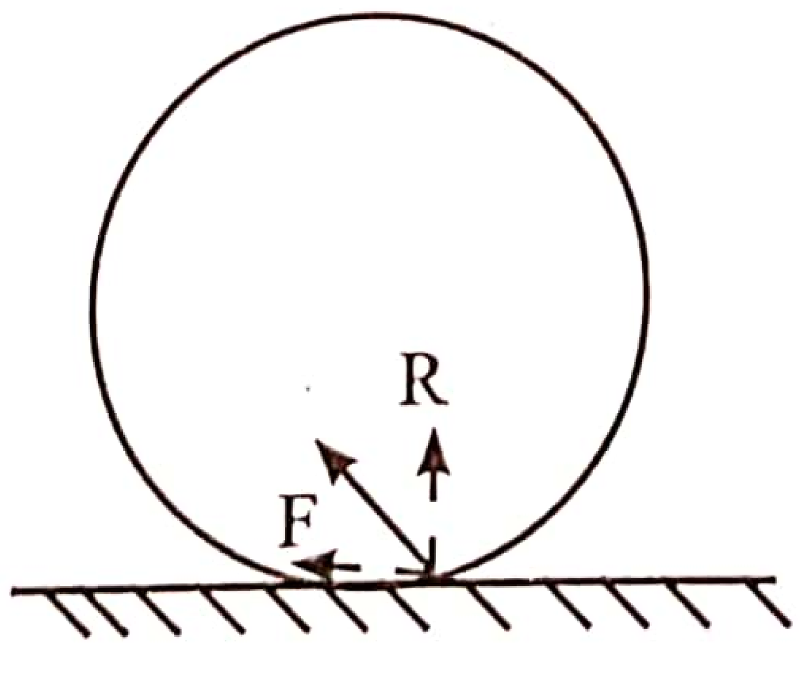

(4)


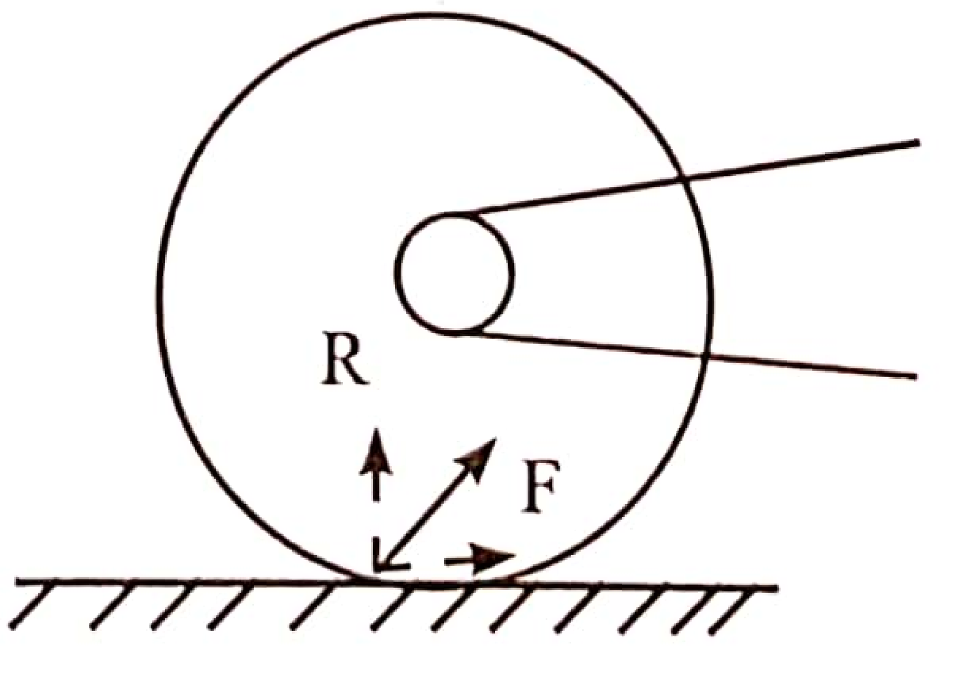
(5)


Such questions have been discussed in previous past papers. The 27th question of paper 2005 is like this question. It has a moving bicycle there. I have given a review of the direction of forces there. Even in the first essay question of paper 2002 a detailed explanation has been given. I will present that description again.

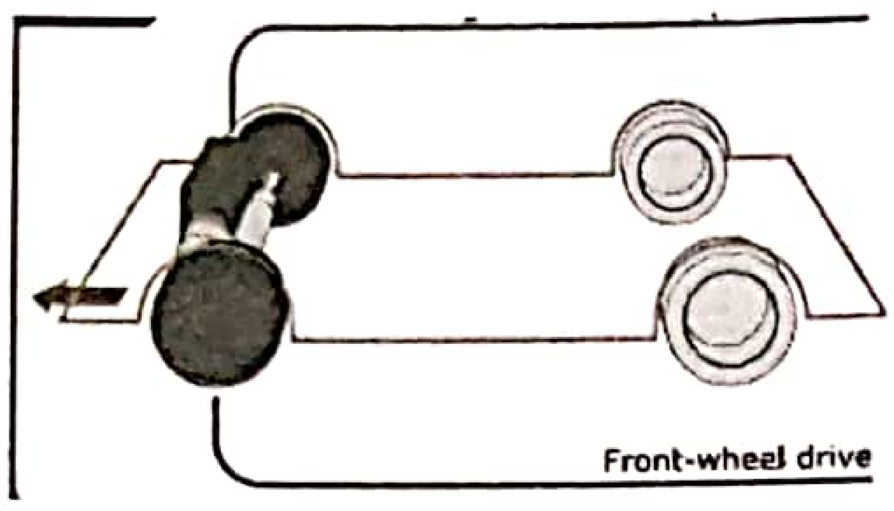
The forces acting on the rear wheel of a bicycle has been shown below in figure 1. There is a driving torque on the rear wheel at such an instance. These are considered as the driven wheels. Such a wheel presses the ground backward. Then keep an eye on the nature of the surface on the ground. The surface shape has been distorted than the normal way to understand easily. Sometimes even if we do not see this for our naked eye, it is true in the limit of atomic scale. Then does not the horizontal force (friction) from the ground on the wheel is acting to the forward direction?

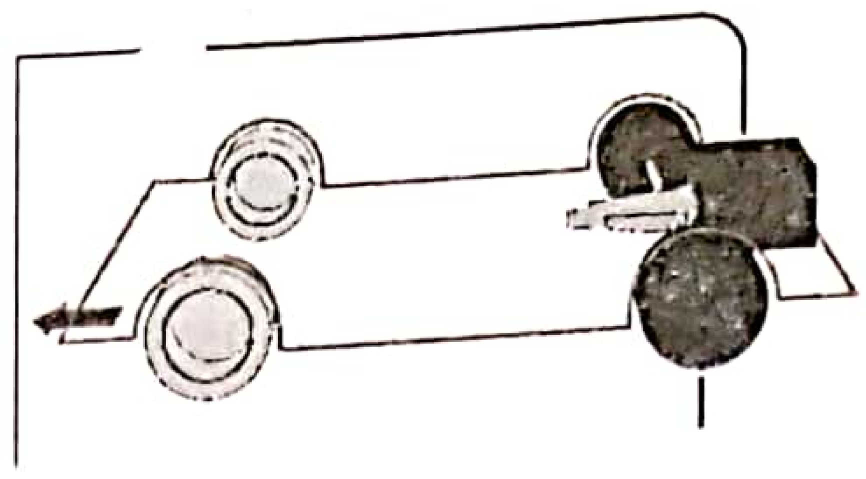
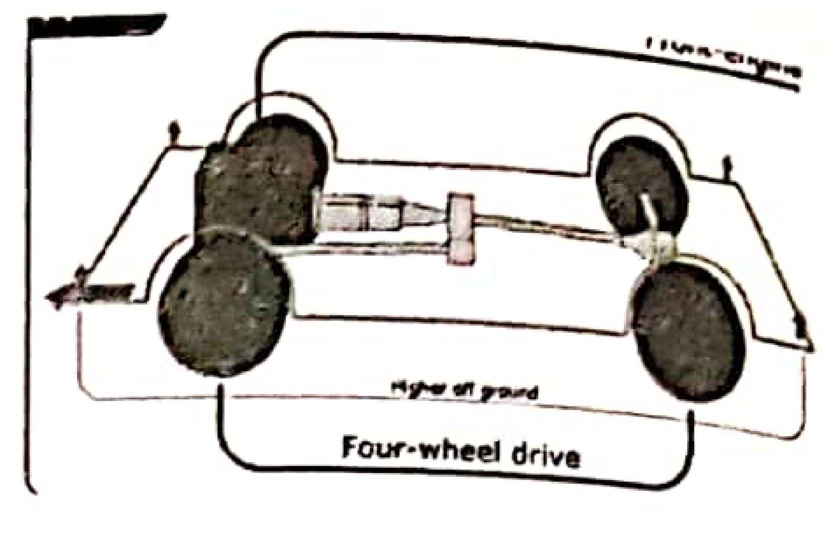
The forces on a rolling undriven wheel from the ground is shown from figure 2. Then the firctional force is acting backwards on the wheel.





Here it is enough to consider the wheels that are driven and undriven. Motor vehicle A is a vehicle which has its front wheels only connected to the engine. Therefore, the force that is needed to go forward is supplied by pressing of front wheels on the ground. Then the rear wheels are not driven wheels. In motor vehicle B, it has rear wheels that are connected to the engine. Then clearly the force to go forward should be obtained by the pressing of rear wheels on the ground. Therefore, the correct figure is (4).



If you think in a simpler way, then from the wheels that are not connected to the engine cannot provide the forces that needs to go forward. It is great and it it is like that. It will be like God is driving.

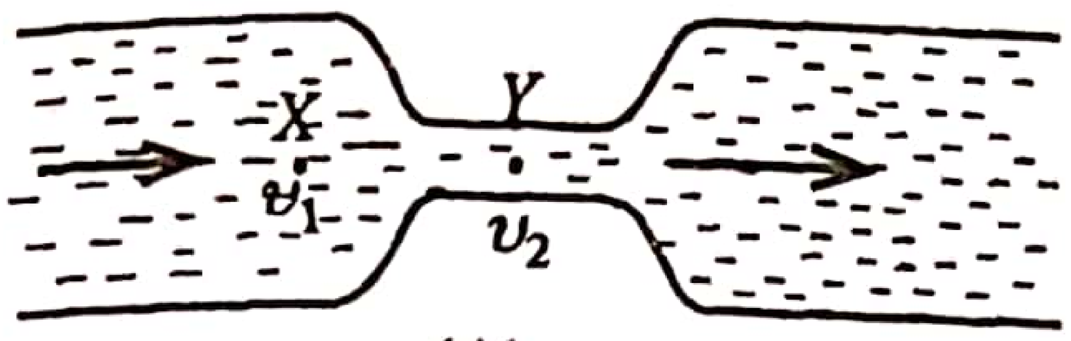
Early vehicles came as rear-wheel. But as the engine is at the front, the prop-shaft should be rotated and transmitted to the rear wheels to get the force that is needed to rotate the rear wheels. The energy is wasted here and the front wheel vehicles were started to manufacture as the rotation of the front wheels of the engine is more efficient and they can be used to save fuel. Now there are four-wheel drive vehicles too. Even though they can go faster, they consume lot of fuel. You should be able to draw the direction of forces acting on the wheeld by the ground in such a vehicle. Different sorts of advantages and disadvantages are shown below.

As the engine and the driven wheels are at the same side of the vehicle in front wheel vehicles, there is more space in the vehicle. The prop-shaft that is going across the middle of the vehicle is not needed here. As the weight of the engine is also acting on the front, by having the driven wheels are at the front, the front part of the vehicle is properly touched with the ground.

But many vehicles which can go faster like racing cars are made as rear wheel drives. The reason for this is the weight of the vehicle is mostly towards the back when the car accelerate faster. Then more connection with the ground/ good traction can be obtained by the rear wheels. In higher accelerations the front wheels tend to rise up. The other advantage of rear wheel vehicles is that the front wheels can be cut well as they are mainly used for the driving/ controling of the vehicle. The front wheels can be turned left or right as you wish compared to front wheel drive vehicles.

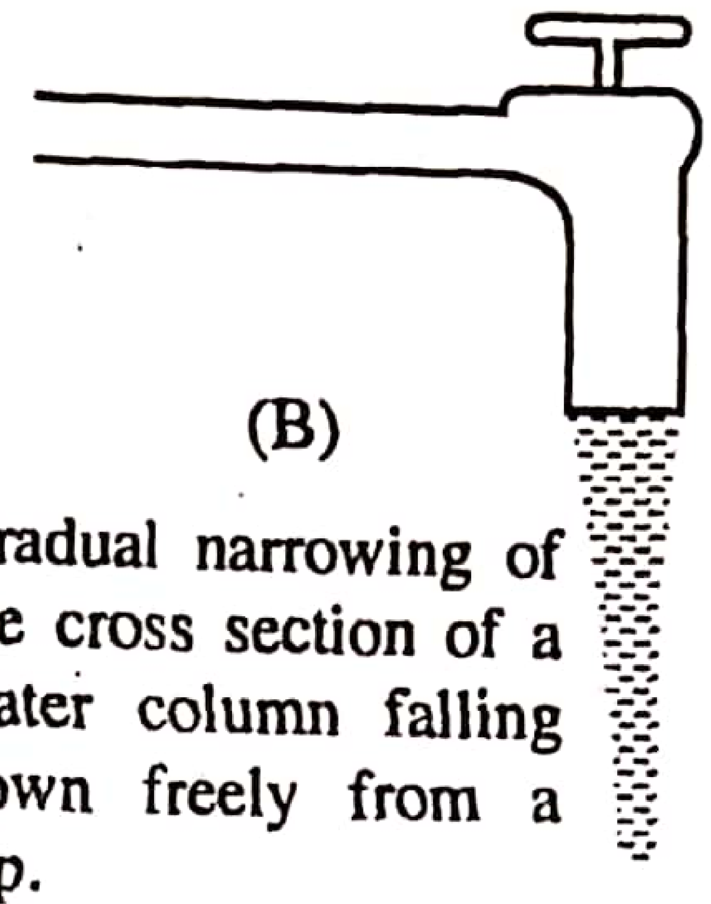
Untill 1970, 80s many vehicles were manufactures as rear wheel drives. Still bicycles and motor cycles are as rear wheel drives. As the front wheel is needed to rotate freely, there canbot be another alternative to be used. Actually, the vehicle does not move forward due to the engine. From the engine, the wheels should be rotated and get the forces by pressing/hugging the ground. Some does not like to call these forces as frictional forces. We always think that frictional forces act against the motion. But this feeling or the concept is wrong. Frictional force is the force that is acting on a surface or the surface that is obtained by pressing the ground or parallel to the ground. According to the pressing direction, it can be towards the direction of the motion or it can be against the direction of the motion.

39. Consider the following physical phenomena.



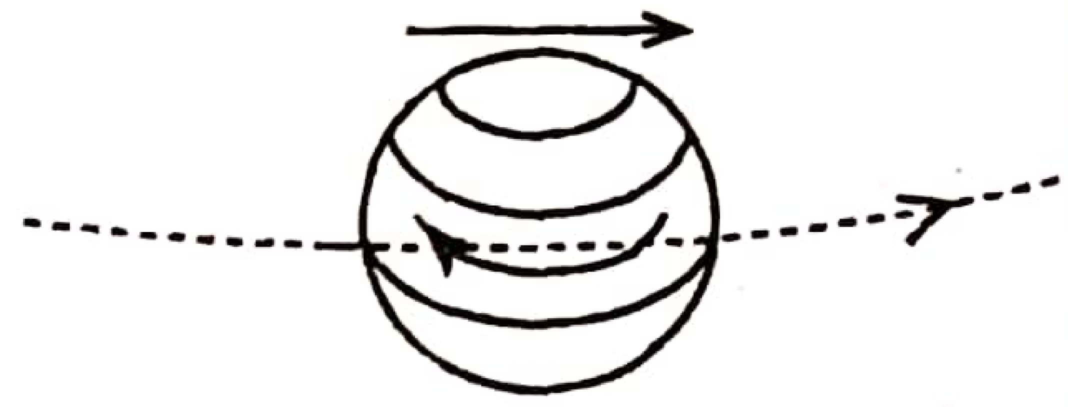
(A)

Water flowing through a tube having two different areas of cross-section; speed of water at Y (v_2) > speed of water at X (v_1).



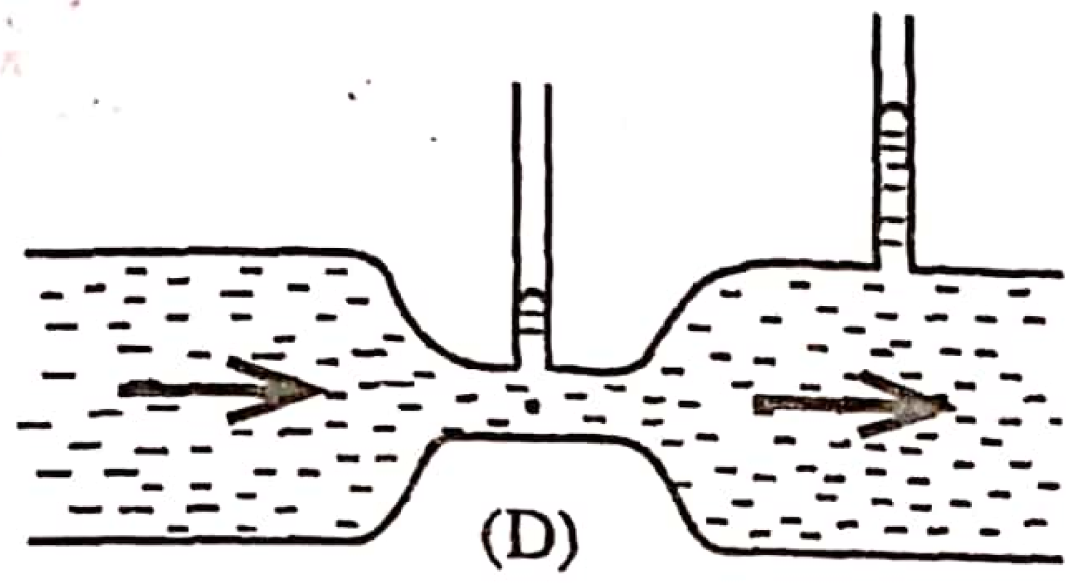
(B)

Gradual narrowing of the cross section of a water column falling down freely from a tap.



(C)

Deflection of a cricket ball which is moving while spinning.

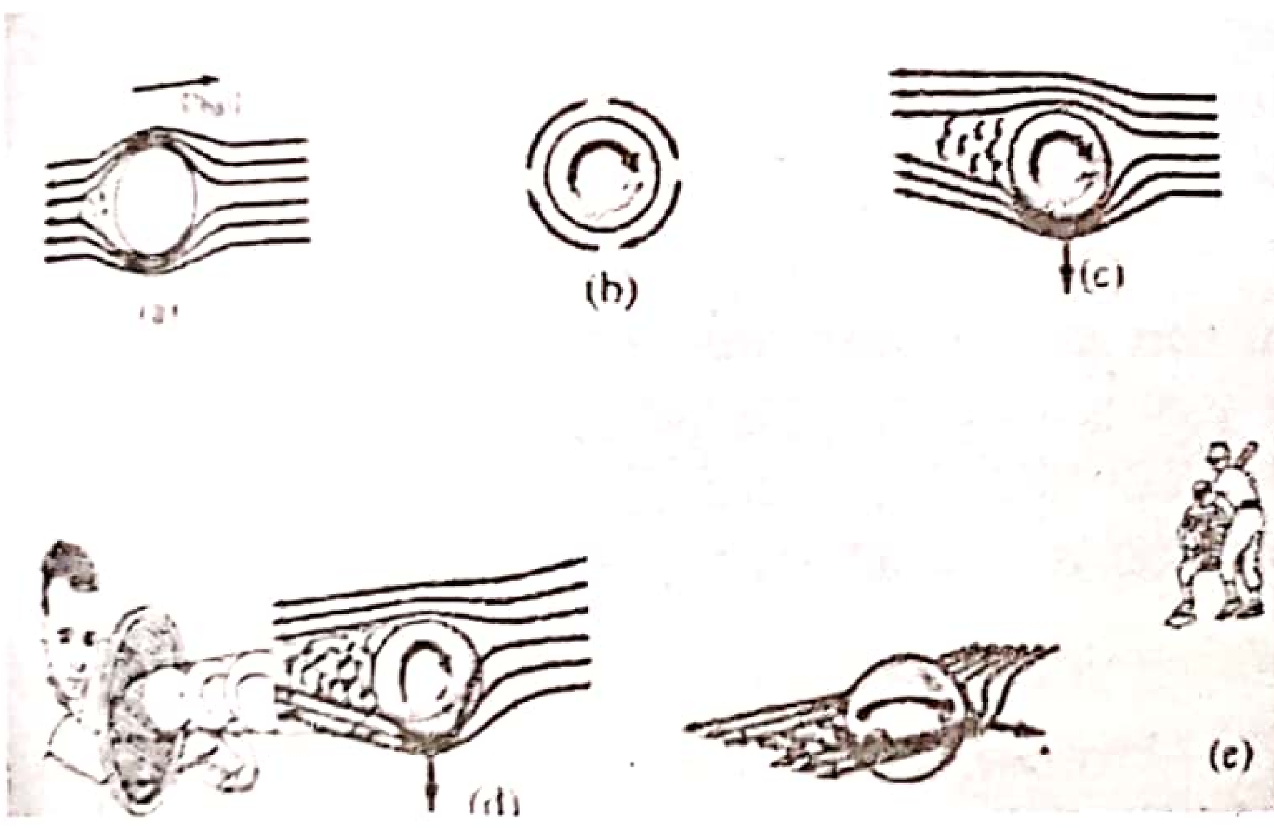


(D)

Existence of a height difference in the liquid columns in vertical tubes.

Which of the above phenomena can be explained using the Bernoulli's theorem?

(1) A and D only. (2) B and D only. (3) C and D only.
(4) B, C and D only. (5) All A, B, C and D



You can get the answer at a glance. To explain (C) and (D) only we need Bernoulli's theorem. To explain the instance of (A) we need the equation of continuity which is $A_1v_1 = A_2v_2$. Even this equation is needed for (D), to explain the height difference of liquid columns in vertical tubes, we need Bernoulli's theorem.

Bernoulli's theorem is not needed to decide that the speed of the water column is increased when falling downwards from the tap. You can just say it. Then according to $A_1v_1 = A_2v_2$, if $v_2 > v_1$ then you can decide that $A_2 < A_1$. If you need to find the speed of a water column that is falling downwards from the tap, you can apply Bernoulli's theorem.

If u is the velocity of a water column of point P which is just outside the tap, then applying Bernoulli's theorem, $\pi + \frac{1}{2}\rho u^2 + 0 = \pi + \frac{1}{2}\rho v^2 - \rho gh$

$$\frac{1}{2}v^2 = \frac{1}{2}u^2 + gh;\ v^2 = u^2 + 2gh$$

You can get the answer if you just apply $v^2 = u^2 + 2gh$. Finding the speed of a liquid column is complex. At above, I have considered the two points that are exposed to air. The pressure of the two points can be considered as the atmospheric pressure in such points. We simply do not know the pressure inside the liquid column. In such occasions, you consider that the liquid column is narrow and the normal tradition is to consider the speed of the points in any horizontal cross section as equal.

We need Bernoulli's theorem to explain the deflection of spinning ball. It is a known fact by everyone. It is better to consider the function of a spinning ball. Consider a ball which moves from left to right in the air (without a spin). The air current is moving from right to left according to figure (a) relative to the ball.

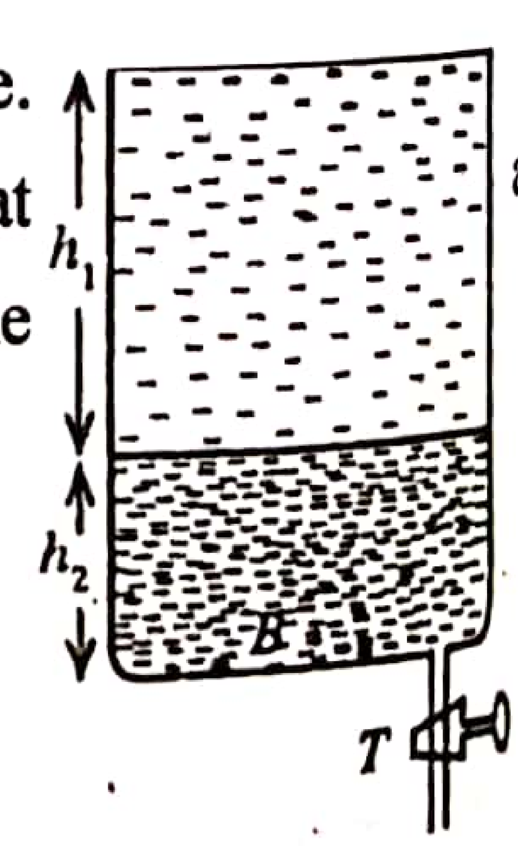
Figure (b) represents a ball with a top spin. Top spin is a ball that spins to the given direction (clockwise) in a horizontal axis across its centre which is perpendicular to the paper plane. Due to the ball and the air friction and due to the viscosity of the air, the gas layers near to the ball surface are drawn towards the direction of the ball spin. Due to this, the speed of the gas layers on the bottom of the ball is higher compared to the gas layers on the top of the ball relative to the ball. Why? If there is no spin, the gas flew from right to left relative to the ball. Due to the spin, the bottom gas layers closer to the ball also spins it adds to the speed whereas the speed of the top gas layer of the ball get reduced due to the spin. Why? The speed that is obtained by the gas layers due to spin on top of the ball is towards the opposite direction of the gas current as in figure (a).

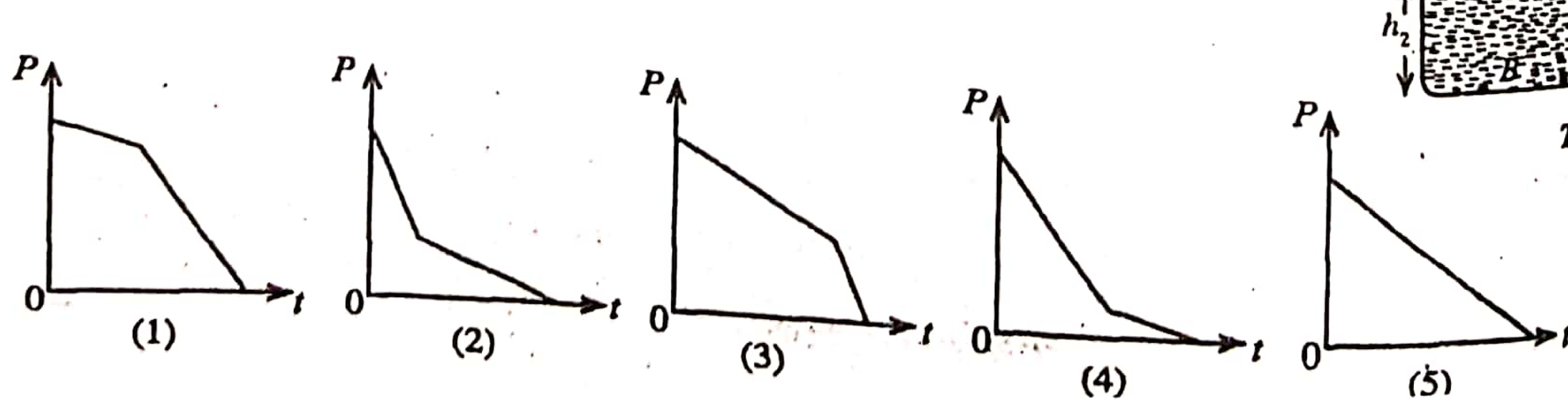
The asymmetry occurs due to this has been shown in figure (c). The flow lines at the bottom of the ball are closer relative to the ball (speed is higher). Likewise, the flow lines on the top of the ball are separated (speed is less relatively). Therefore, according to Bernoulli's theorem, the air pressure on top of the ball is increased relative to the bottom. So, the ball is deflected downwards. To keep the serve shot on the field, this top spin is being used in the game of tennis (figure d). Even in cricket it is being used. A top spin ball will not deflect on both sides and tries to deflect down and go forward.

If we look from the top to the bottom across the vertical axis of the ball and if the rotaion is clockwise, then what is happened is shown in figure (e). The ball will then deflects to the shown direction (to the right side if looked from the top). It is an off spin ball to a righ handed batsman. If we look from the top to the bottom across the vertical axis of the ball and if the rotaion is anticlockwise, then you will understand it easily as a leg spin ball. The ball will deflect on a horizontal plane. It will be towards the left side if it is looked from top to bottom. Such tactics are being used by the baller in the game of baseball. The ball path can be curved towards the batsman or away from him.

In golf, the ball is given a back spin. In a back spin, the ball is spinned to the opposite direction which is shown in figure (b). Then you can easily decide that the ball tries to deflect upwards. It is lift off. A ball hit with a back spin remains more time in the air than a ball which is not hit with a back spin. This is an important factor is golf. If the surface of the ball is very smooth, the above mentioned deflections cannot be obtained easily. When the ball is spinned, there should be the needed friction which should be drawn by the near by air layers with ball. People are relcutant to put a spin ball from the new ball due to this reason. You can identify that why there are dimples like acne in golf balls. Golf balls are not rubbed on the trousers. There is no use even if you do so. As there is a seam in the cricket ball, it is complex to study the deflections. Even seam also deflects after hitting the field. The correct answer is (C) and (D).

40. A cylinder contains two immiscible liquids filled to heights h1 and h2 as shown in figure. If the tap T at the bottom Is opened at time t = 0 and the liquids are taken out slowly at constant volume rate, the variation of pressure (P) due to liquids at the point B at the bottom of the cylinder with time (t) is best represented by





Hydrstatics

There is no need to write equations for this question. Even there is no inability in writing equations. But you can save the time by getting the answer from the logic. However, P should be reduced with time. There are no graphs that show P is increased with time. There are no curves also. As there are two liquids with different densities, there cannot be a straight line with same gradient. There should be some difference. There should be a sort of division. From this logic, the choice (5) can be removed.

In the cylinder, there is more liquid with the height of h_1. As the liquids flow at an equal volume rate, the liquid with the height of h_2 will be removed and liquid with the height of h_1 will be removed afterwards. The liquid with the height of h_2 should take more time for the complete removal. These are simple logics. From this logic, you can remove (3) and (4) directly. In both of their shapes, the time taken for the second liquid removal is less compared to the first liquid removal. This cannot happen.

Now, how can you select the correct choice from (1) and (2)? From simple logic this can be done too. At t =0, the total pressure is obtained from both liquids. In the graph of (2), if the second straight line is drawn backwards till t =0, the point where it cuts the pressure axis is corresponding to the pressure only from the second liquid with a height of h_1 is filled to a height of (h_1+h_2) in the cylinder. That means $(h_1+h_2)\,\rho_1 g$. But when t=0, the pressure of point B is $(h_1\,\rho_1 g + h_2\,\rho_2 g)$. As $\rho_2>\rho_1$, then $h_2\,\rho_2 > h_2\,\rho_1$. Therefore, when the second straight line is drawn backwards, the point where it cuts P axis cannot be at a higher place compared to the initial place with both liquids. From this logic the first graph can be removed. The correct one is (2).

If you write equations, then consider V as the volume flow rate of the liquid which goes out in time t. The reduced height of the liquid is Vt. If the cross-sectional area of the cylinder is A, then the reduced height in time t is Vt/A. So, pressure P of point B when the liquid at the bottom is going out will be,

$$P = \left(h_2 - \frac{Vt}{A}\right)\rho_2 g + h_1\rho_1 g; \quad P = -V/A\,.t\,\rho_2 g + h_1\,\rho_1 g + h_2\,\rho_2 g$$

The graph of t against P should be a straight line with a neagative gradient. The gradient is proportional to ρ_2. Likewise, you can decide intelligently that when the bottom liquid is flown out, the gradient of the relevant variation should be proportional to ρ_1. As $\rho_2>\rho_1$, the gradient of the first straight line should be greater than the second line. It is proved by (2).

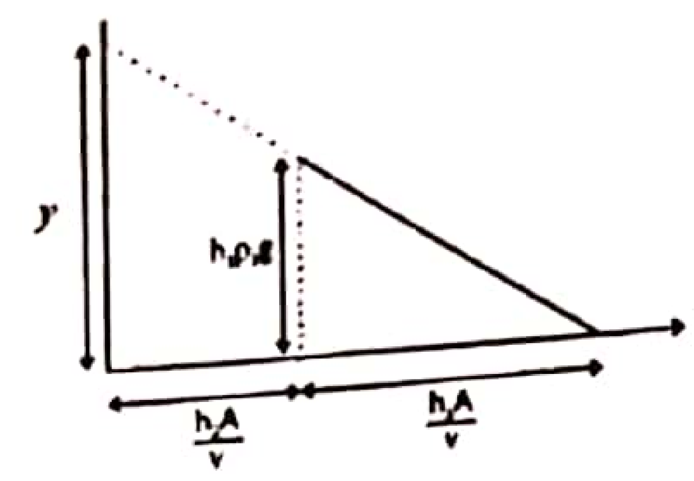
You can analyze by writing more and more equations. If t_2 is the time taken by the bottom liquid, then once it is completely removed, the pressure is obtained by the liquid only on the top. When

$$t = t_2,\ P = h_1\rho_1 g;\ h_1\rho_1 g = -V/A\,.t_2\rho_2 g + h_1\rho_1 g + h_2\rho_2 g$$

$$t_2 = \frac{h_2 A}{V}$$

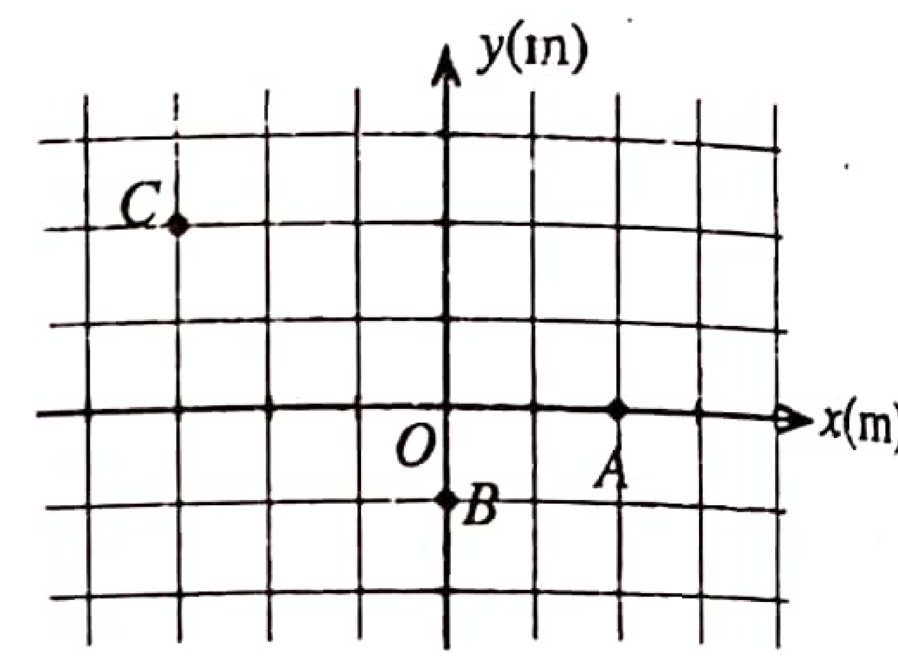
This can be just taken. The total volume of the bottom liquid is h_2A. As the liquid is removed at a uniform rate of V, the time taken to remove a volume of should be $\frac{h_2 A}{V}$ Likewise, time taken to remove a volume of h_1A is t_1 whereas $t_1 = \frac{h_1 A}{V}$. As $h_1 > h_2$, $t_1 > t_2$. From this, the first logic is proved. If needed you can see that my second logic is also correct. $\frac{y}{\frac{h_2 A}{V}+\frac{h_1 A}{V}} = \frac{h_1\rho_1 g}{\frac{h_1 A}{V}};\ y = (h_1+h_2)\rho_1 g$

All these arguments can be done as the liquid is going out at a constant volume rate. To do this practically, the tap should be adjusted when the liquid flows. When the liquid flows, the tap also should be opend bit by bit. Otherwise if the tap is kept opened at a position, initially the water flows quickly. Later on as the pressure gets reduced, the liquid flow rate is also get reduced.



Even though I write this much, from my first logic of time, you can find the correct choice as (1) or (2). Next, you can just think that the gradients of the straight lines should be proportional to the densities of the liquids. What else will it be proportional? Even from that blind sight, you can think that the correct variation is (2). Actually, when a liquid flows, the pressure is not given by hρg. This is only the static pressure. But the examiners have given the point B far away from the liquid flow. The flow lines are prohibited. If the point B is marked near the tap, this question cannot be done.

41. **A small object is initially at rest at point 0, and due to an Internal explosion it breaks into three parts and move away. At a certain instant after the explosion, the location of three moving parts are shown by the points A, B and C in figure. If the mass of the part which is al point A is 6 grams, what is the mass of the object (in grams) before explosion?**

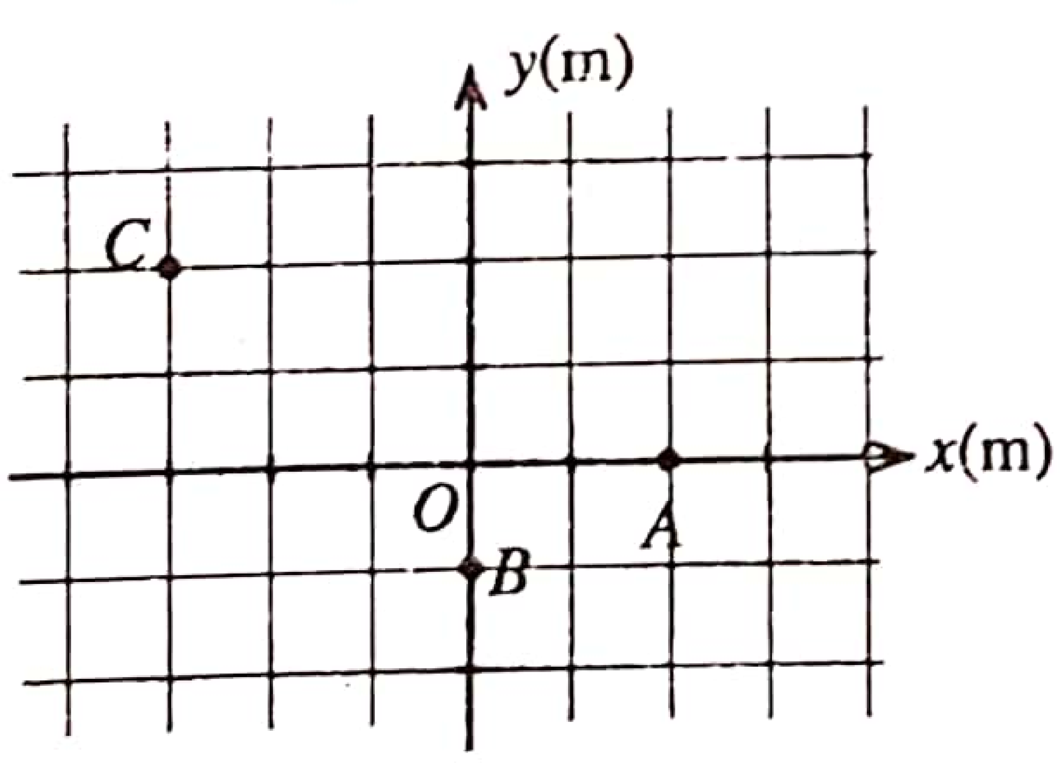


(1) 6 (2) 9

(3) 12 (4) 15

(5) 18

As soon as you see an explosion, you should remember conservation of linear momentum. Count the squares to fond the velocities of the pieces. Do not think far than that. Initially the momentum of the object is zero. Therefore, the vector sum of the momentum of the pieces after the explosion should also need to be a zero. According to the way that the question is asked, the motion happens in the plane of x-y. Therefore, the net vector momentum across x and y direction have to be zero. The velocity of piece A is 2 units to +X direction (2 squares), The velocity of piece B is one unit to -Y direction.



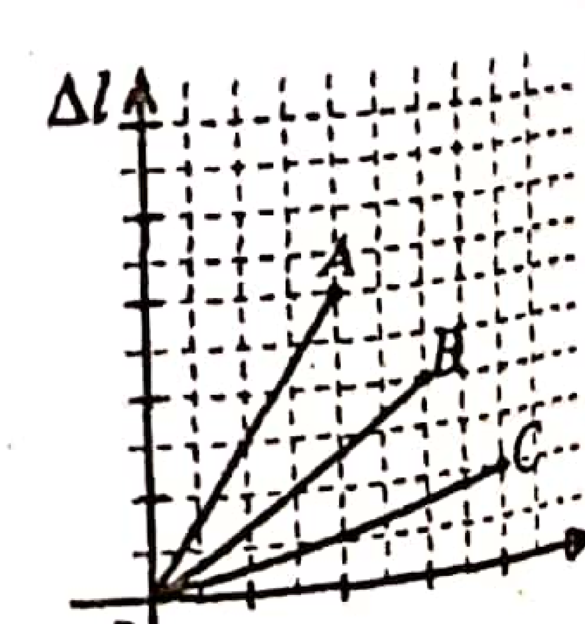
The velocity of piece C is 3 units to -X direction and 2 units to +Y direction. Now use conservation of linear momentum.

$$\rightarrow 6 \times 2 = m_C \times 3;\ m_c = 4$$

$$\uparrow 4 \times 2 = m_B \times 1;\ m_B = 8$$

Therefore total = 6 + 8 + 4 = 18 g

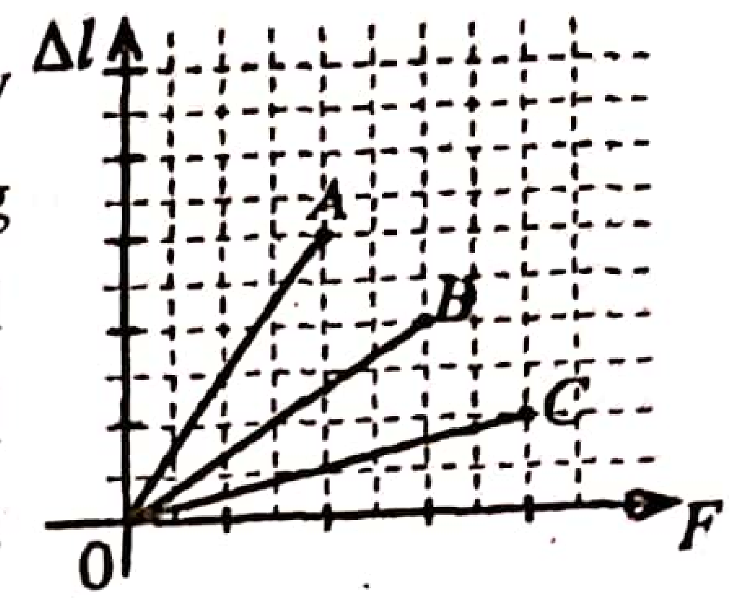
42. **Figure shows the variation of the extensions (Δl) produced by three different metal rods A, B and C with the force when they are subjected to a tensile force F. If E_A, E_B and E_C are the corresponding energies stored in the rods due to extensions, then,**



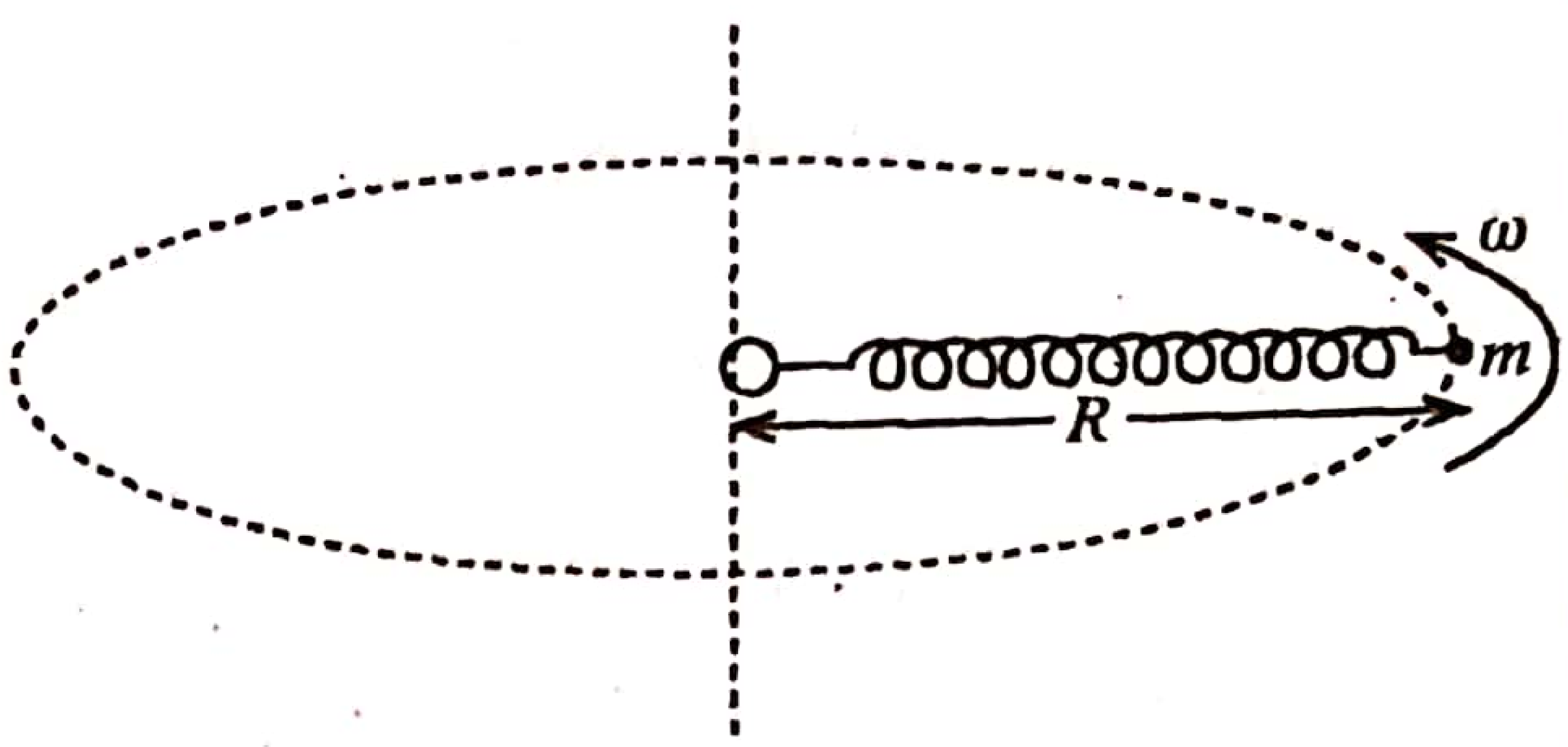
(1) $E_A > E_B = E_C$ (2) $E_A = E_B > E_C$ (3) $E_A = E_B = E_C$

(4) $E_A > E_B > E_C$ (5) $E_A < E_B < E_C$

This is very simple. It is not worth for the 42nd question. The stored energy can be obtained by ½FΔl. Estimate the areas of the triangles that are surrounded by the relevant three straight lines and the axis of F. This is a question of counting squares. There are three such questions in this paper. Forget about ½ as you need to do a comparison here. In the vertical axis of A there are 3 squares of horizontal lines and 2 horizontal squares. 3X 2 = 6. In B there are 2 marked vertical squares and 3 horizontal squares. 3 X 2 = 6. In C, there are 4 horizontal and one vertical squares. 4 X 1 = 4. Therefore, $E_A = E_B > E_C$. If needed, you can count the small squares too.



43. A light spiral spring has an unstretched length l and a spring constant k. A small object of mass m is attached to one end of the spring and the system is rotated about a vertical axis that passes through a small light ring attached to the other end of the spring as shown in figure. If the object travels along a circular path of radius R with constant angular speed ω, keeping the spring on a horizontal plane, then



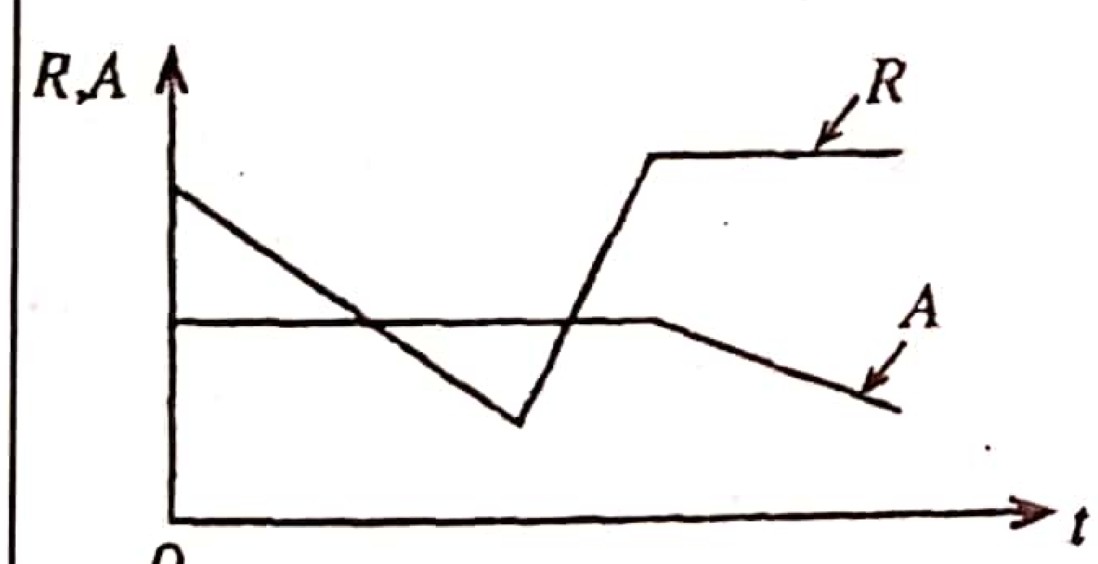
(1) $\omega = \sqrt{\frac{k}{m}\left(\frac{R-l}{R}\right)}$ (2) $\omega = \sqrt{\frac{k}{m}}$ (3) $\omega = \sqrt{\frac{k}{m}\cdot\frac{l}{R}}$

(4) $\omega = \sqrt{\frac{k}{m}\left(1-\frac{R}{l}\right)}$ (5) $\omega = \sqrt{\frac{k}{m}\cdot\frac{R}{l}}$

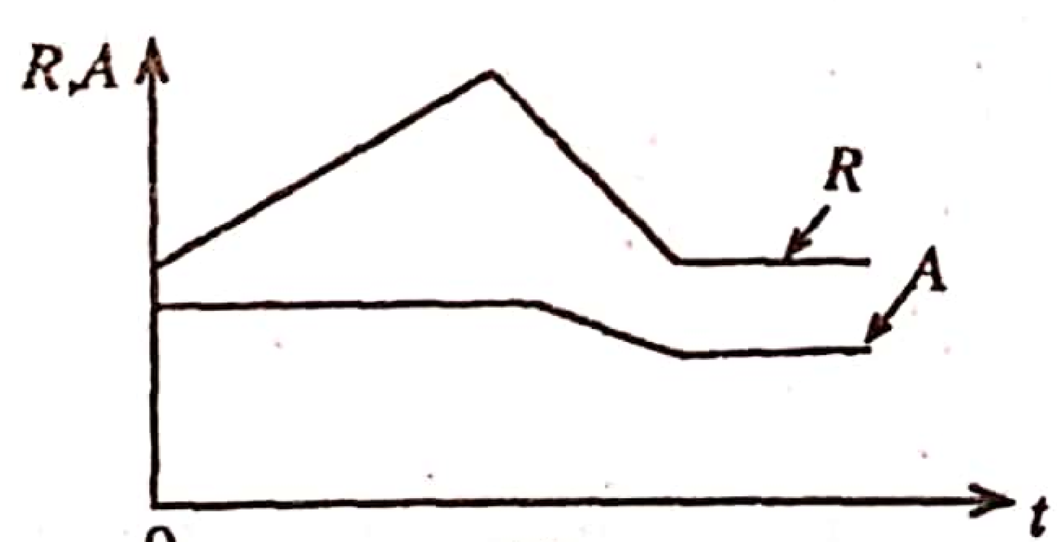
Circular Motion **02**

This is also very simple. The centripetal force should be equal to mRω2. Even this can be given to the first 20 questions. The horizontal force on m due to the spring = k (R-l) {k X extension} $mR\omega^2 = k(R-l)$. The answer is in your hand. All you need to write is this in your rough sheet.

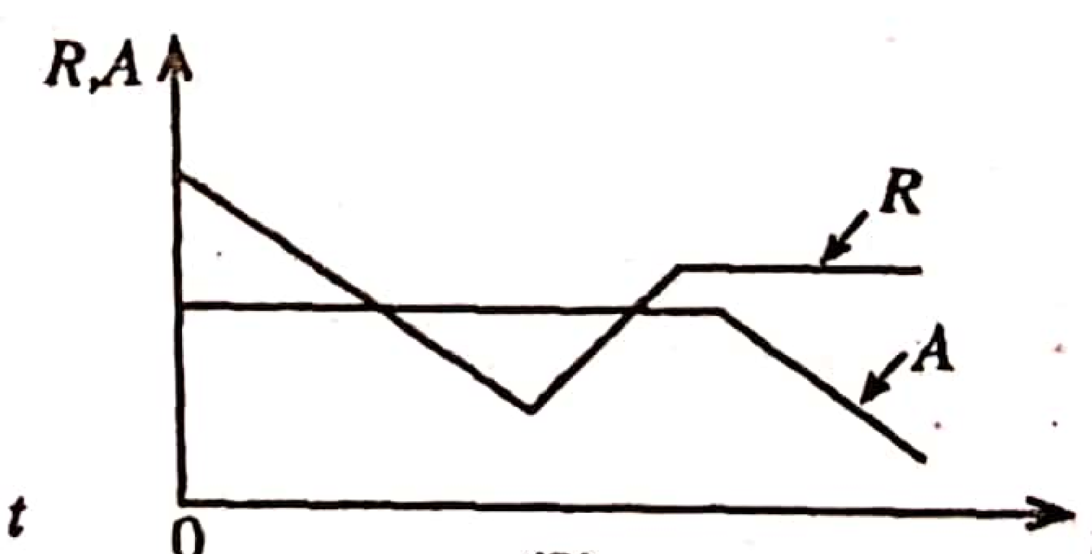
44. A certain volume of air, isolated from the atmosphere at 30 °C, is first heated up to 80 °C and then cooled down to 15 °C at uniform rates. Both heating and cooling are done at constant pressure. Dew point of the isolated air is 25 °C. The variations of relative humidity (R) and absolute humidity (A) of the air volume with time (t) are best represented by



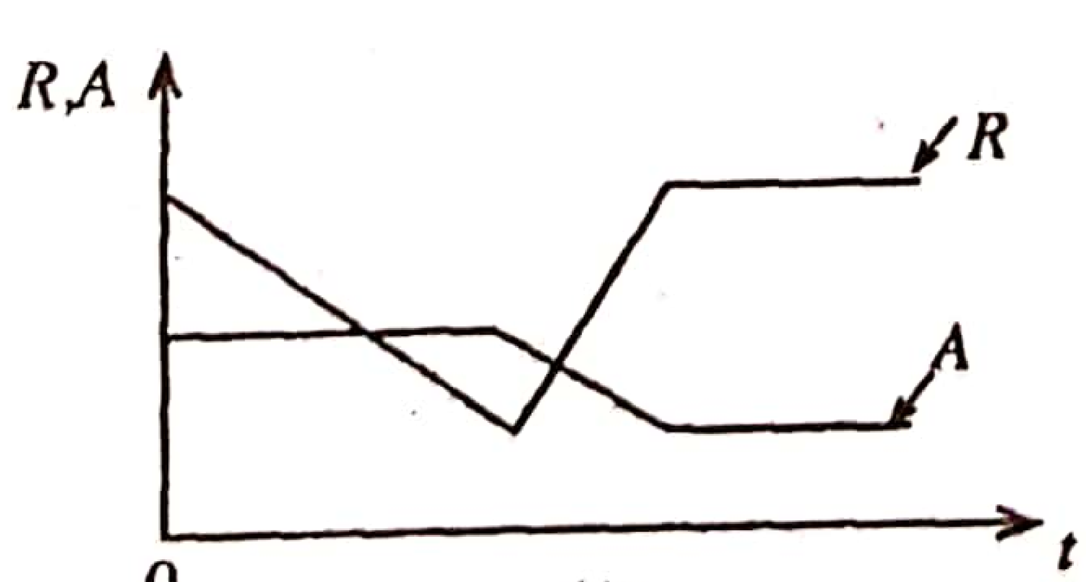
(1)



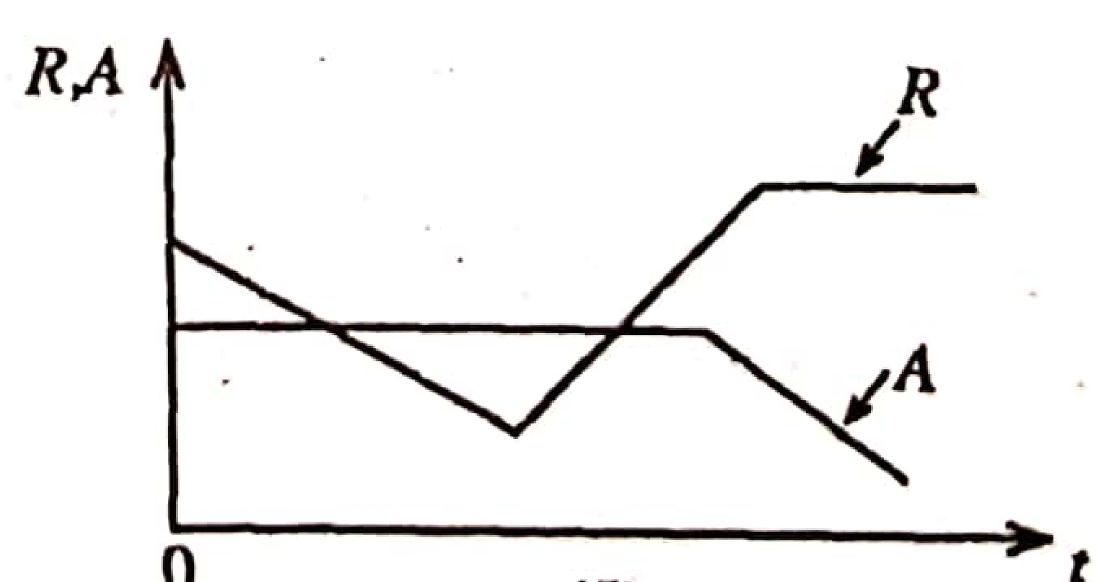
(2)



(3)



(4)



(5)

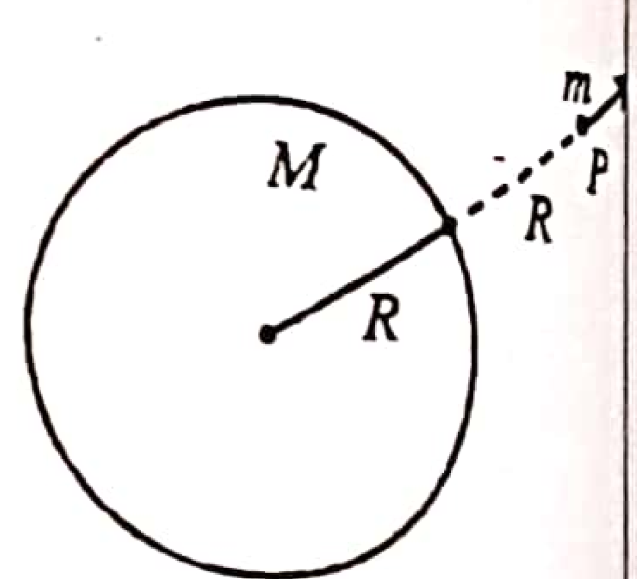
Hygrometry **04**

The relative humidity is remained as a constant when it is cooled than the dew point. In the dew point the relative humidity is 100%. What needs to be happened after that? When it is cooled than the dew point, the dews are formed and the absolute humidity, where the water vapour mass of a unit volume gets reduced. This point has been discussed many times in the previous past papers. The relative humidity gets 100% and remains to start a constant value at the dew point. Therefore, these two processes start at the same time (at the same place). Such a variation is shown in (1). Even form this point, you can find the answer very quickly.

If you look further, when heated the relative humidity gets reduced (from 30°C to 80°C). Then the temperature is reduced again, the relative humidity increases. When the temperature is reduced from 80°C, it again passes 30°C. The dew point is 25°C. From 30°C to 25°C the relative humidity gets further reduced. Therefore, the initial value at 30°C should have been passed here. It gets saturated at 25°C after sometime. The value of R is constant from 25°C to 15°C (100%). Likewise, till dew point, A is not changed. After that it gets reduced.

There is an extra sentence in the question. 'The heating and the cooling are being done under constant pressure'. Due to this there was an issue for this question. When a certain gas volume is taken and the temperature is reduced under constant pressure, its volume changes. It is like a container with a freely moving piston. If the volume is changed, then the absolute humidity is not in a constant value. When the volume is increased, then the absolute humidity gets reduced. Therefore, it was considered to give ALL correct for this question.

45. A particle of mass m is projected vertically upwards from a point P, which is at a distance of $2R$ from the centre of a spherical planet having a mass M and radius R as shown in figure. The escape velocity for this projectile is



(1) $v = \sqrt{\frac{GM}{R}}$ (2) $v = \sqrt{\frac{2GM}{R}}$ (3) $v = \sqrt{\frac{2Gm}{R}}$

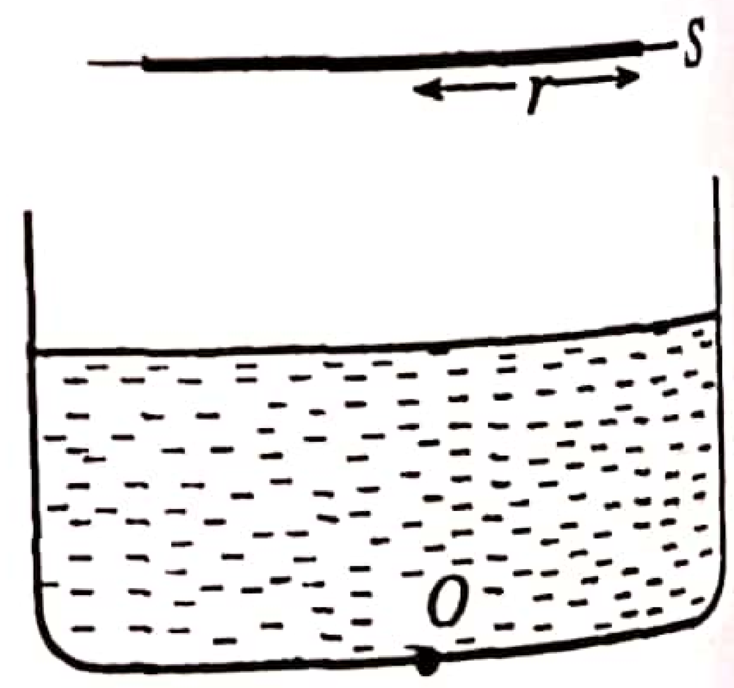
(4) $v = \sqrt{\frac{GM}{2R}}$ (5) $v = 2\sqrt{\frac{GM}{R}}$

Gravitational Force Field

It is an easy question of peanuts. It can be given to the first 10 questions. The escape velocity of a planet surface has been asked many times. Here it has been asked about the escape velocity at a distance of 2R from the centre of the planet. The logic does not change. The escape velocity is the minimum velocity needs to project an object just to a very far way distance (infinity). According to the conservation of energy,

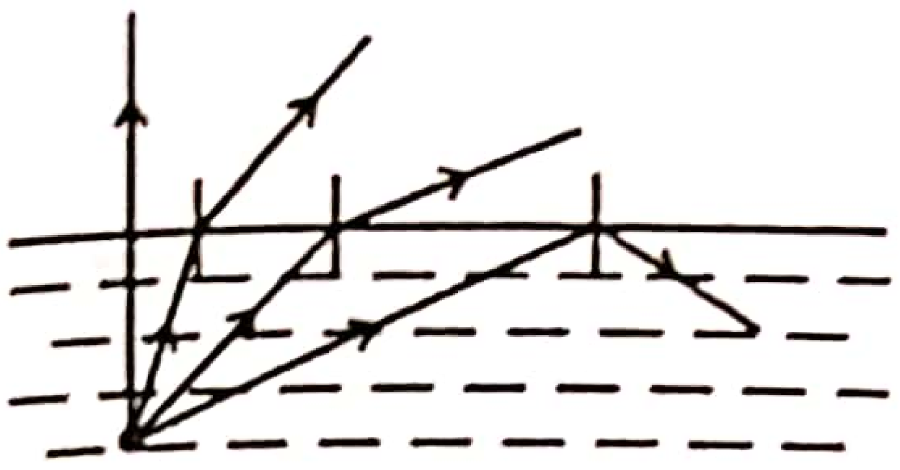
$- GMm/2R + \frac{1}{2} mv^2 = 0; \; v = \sqrt{\frac{GM}{R}}$

46. A point source of light O situated at the bottom of a water tank produces a circular patch of light of radius r on a horizontal screen S as shown in figure. C is the critical angle for the water-air interface. If the light source is moved vertically up by a distance d, the radius of the light patch will

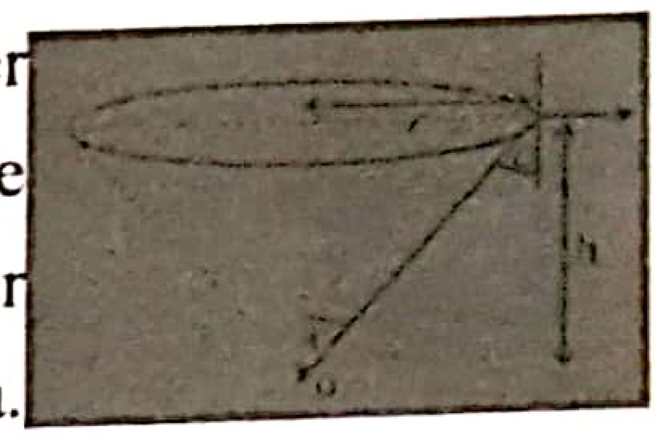


(1) increase to $r+d \sin C$.
(2) increase to $r+d \tan C$.
(3) remain unchanged.
(4) decrease to $r-d \sin C$.
(5) decrease to $r-d \tan C$.

Refraction

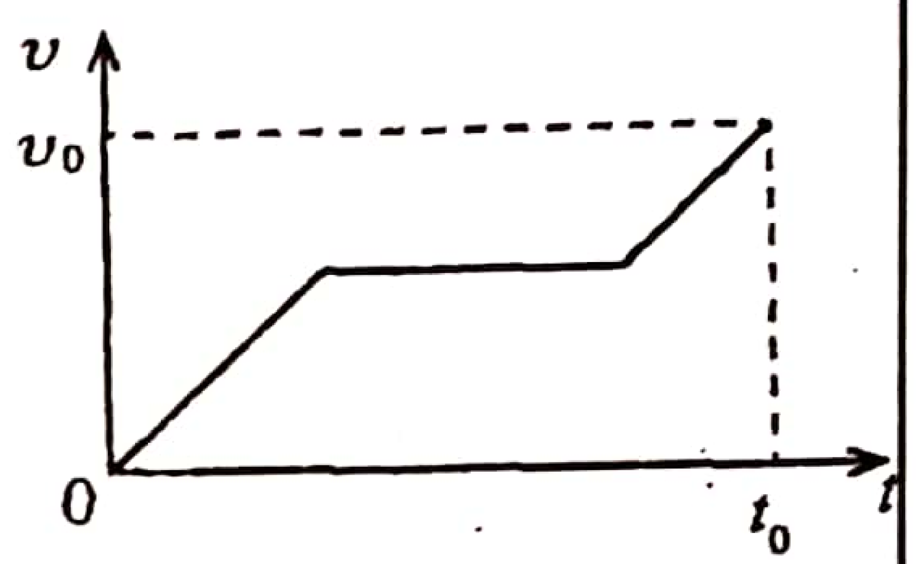


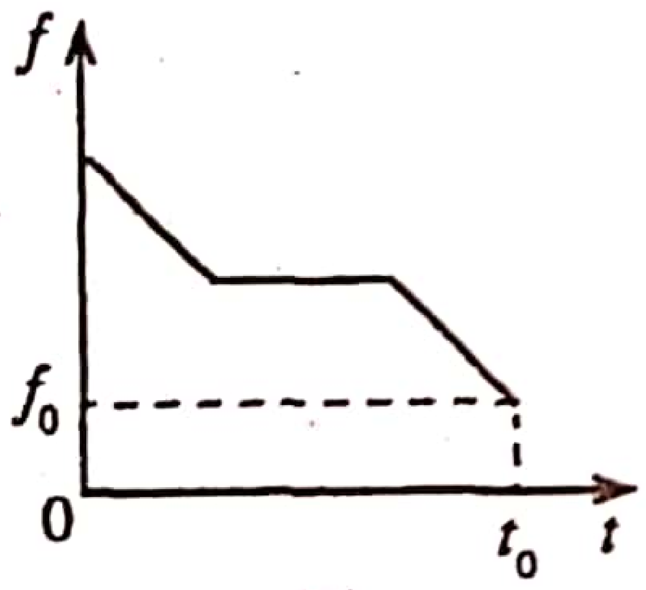
There will not be a light spot as mentioned in the question. When light is refracted from the water surface, the refracted angle will differ in a range from 0^0 to 89.999^0 even if it is subjected to total internal reflection at a certain incident angle. Therefore, the total area outside the water will be illuminated. The area outside water will be illuminated when a light source is kept in a water filled pool.



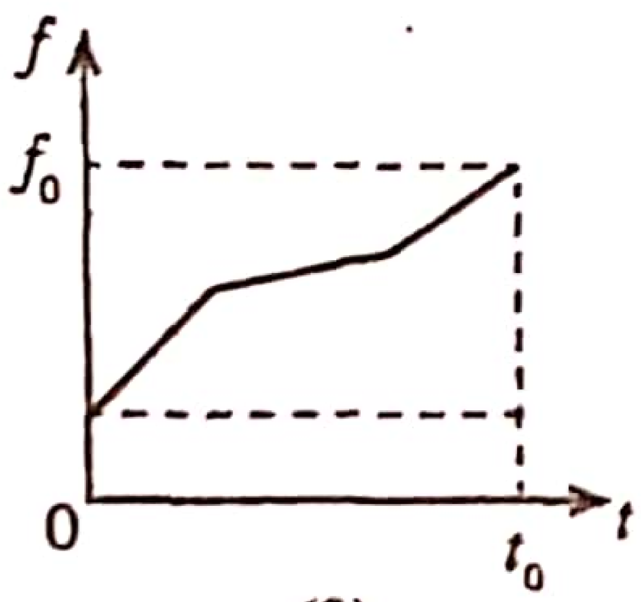
So, it was marked as an ALL. When an opaque disk with a radius r is kept on the water surface, you can allow the source to disappear for an observer who looks from the top. Likewise, for a fish who stays at O in a reservoir with clear water sees the water surface (with sunlight) at day time as a dark surface outside a bright circular area.

47. An ambulance which emits a sound of frequency f_0 from its siren is travelling with constant velocity v_0 along a straight road. A car starting from rest is moving behind the ambulance in the same direction, and the velocity-time graph of the car is shown in figure. The car approaches the velocity v_0 of the ambulance at time t_0. The variation of the frequency (f) of the siren sound heard by a passenger in the car with time (t) is best represented by,

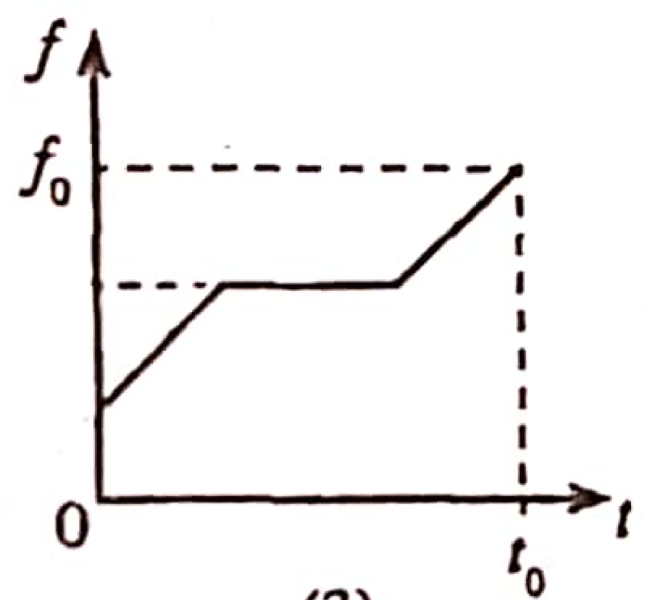




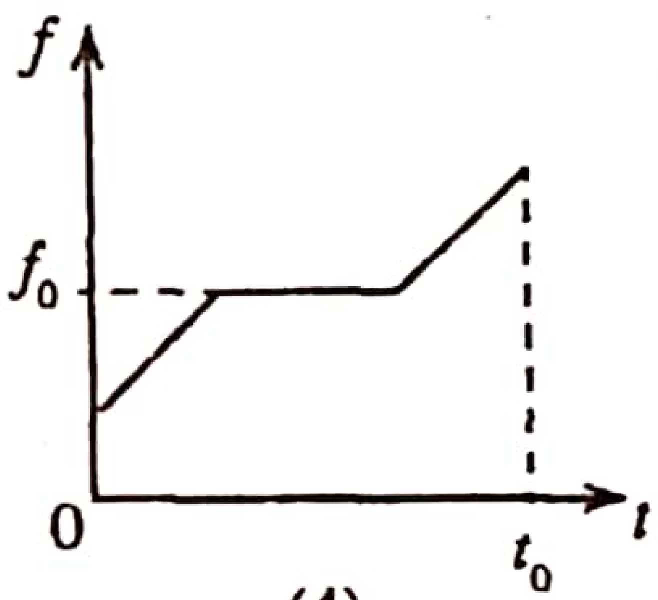
(1)



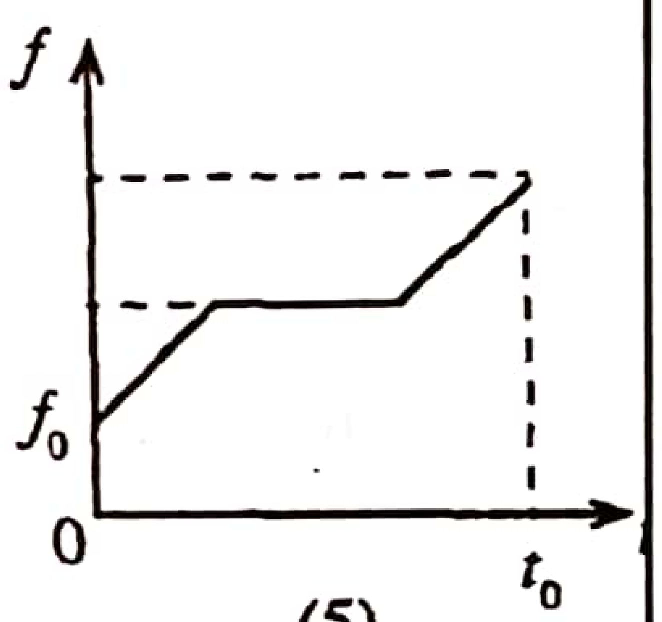
(2)



(3)



(4)



(5)

It is very simple. The motor vehicle is equal to the velocity of the ambulance in t_o time. Then the relative velocity of the ambulance and the motor vehicle is zero. Therefore, the apparent frequency should be equal to f_o in t_o time. Such a representation is shown in (1), (2) and (3).

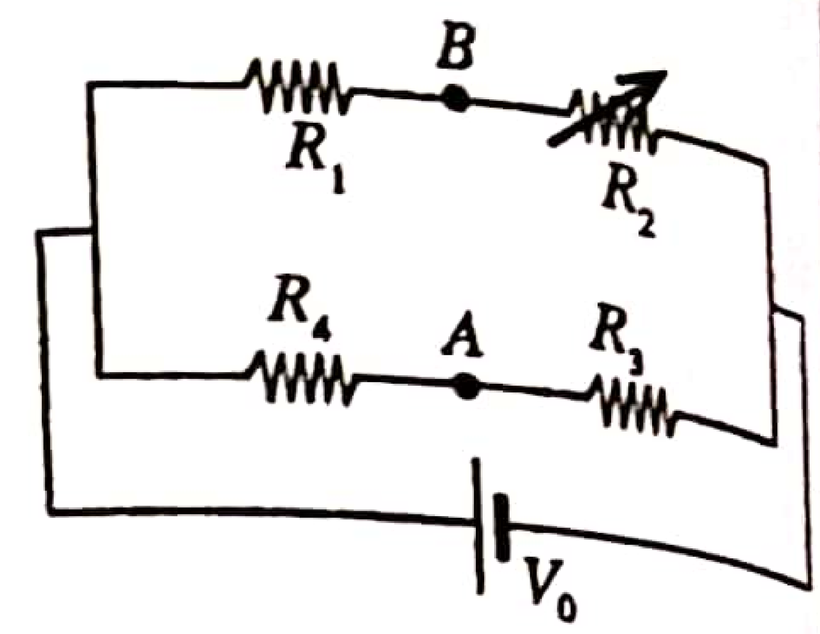
Such problems should be solved with relative velocity without putting equations. To decide the relative velocity, keep the ambulance at rest. The equal and opposite velocity should be given to the motor vehicle too.

$v_0 \longleftarrow\bullet\longrightarrow v$ motor vehicle

$v_0 \longleftarrow\bullet\longrightarrow v$ ambulance

Initially, as the velocity of the motor vehicle is gradually increased from zero, the velocity of the motor vehicle relative to the ambulance is $\leftarrow (v_o - v)$. That means the motor vehicle is like going away from the ambulance. So, the apparent frequency should be less than f_o from the beginning. From that (1) is removed. Out of (2) and (3), you can quickly decide that the correct variation is (3). When the velocity of the motor vehicle is constant, the relative velocity also remains constant. Then the apparent frequency should also be constant ($<f_o$).

48. When the resistance R_2 in the circuit shown in figure is varied from zero to infinity, the potential at A relative to B will change from



(1) zero to zero

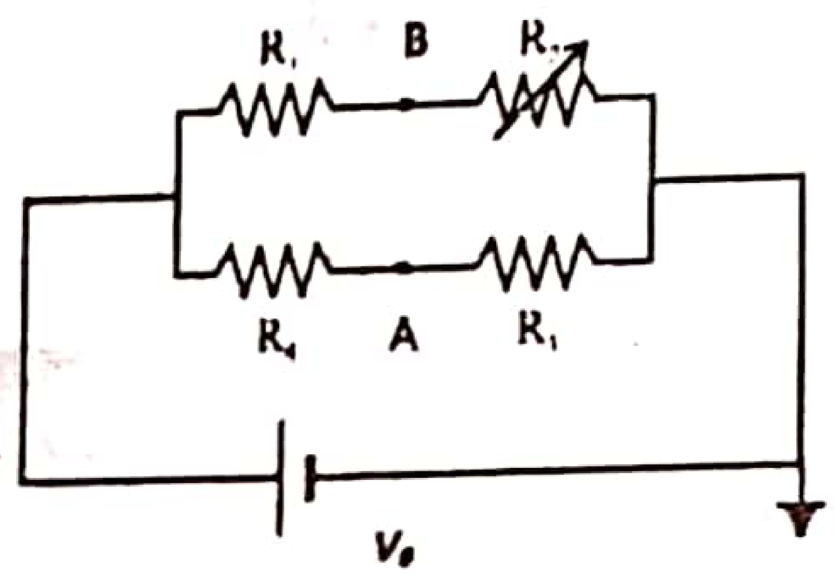
(2) $\frac{R_1}{R_4+R_1}V_0$ to zero

(3) $\frac{R_1}{R_4+R_1}V_0$ to $\frac{R_1}{R_4+R_1}V_0-V_0$

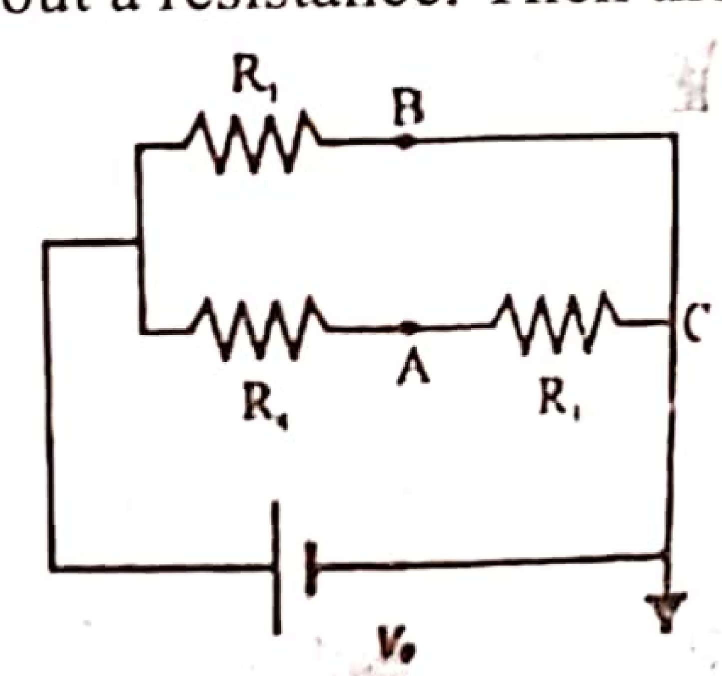
(4) $\frac{R_3}{R_4+R_3}V_0$ to $\frac{R_3}{R_4+R_3}V_0-V_0$

(5) $\frac{R_3}{R_4+R_3}V_0$ to $\frac{R_4}{R_4+R_3}V_0-V_0$

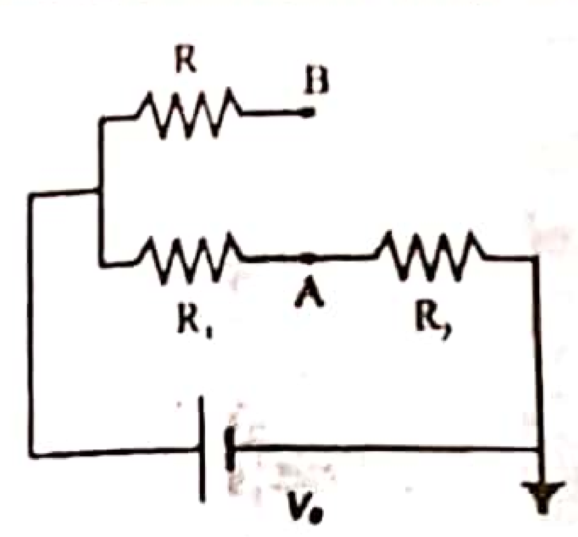
The answer can be obtained easily, by earthing the negative terminal. It is being asked relative to B. When it is asked relative to a point, then earthing one terminal is not an issue. The answer can be obtained even without doing like that way. But the calculations will get longer.



R_2 being zero means connecting a wire without a resistance. Then the circuit will be like this way.



The potential of B is zero here as well as the potential of right side of R_3. $V_C = 0$. Therefore, the potential difference across R_3 is equal to the potential of point A. $V_A - V_C = V_A = R_3 . V_0/(R_3 + R_4)$. If V_0 is across $(R_3 + R_4)$, then how much is across R_3? Now one answer is found. From this you can remove, (1), (2) and (3). R_2 being infinite means that it is being removed. Then the circuit will look like this way.

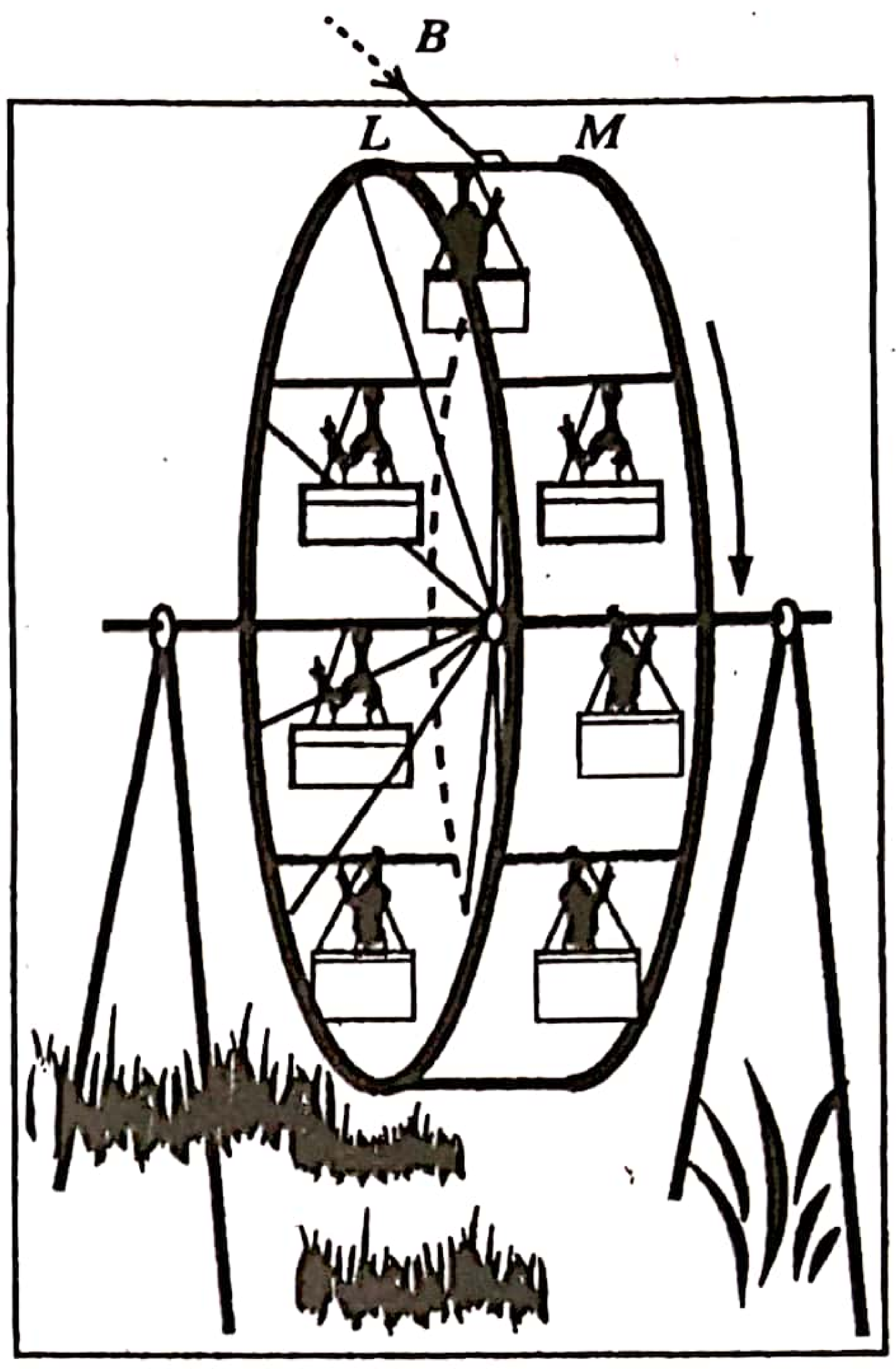


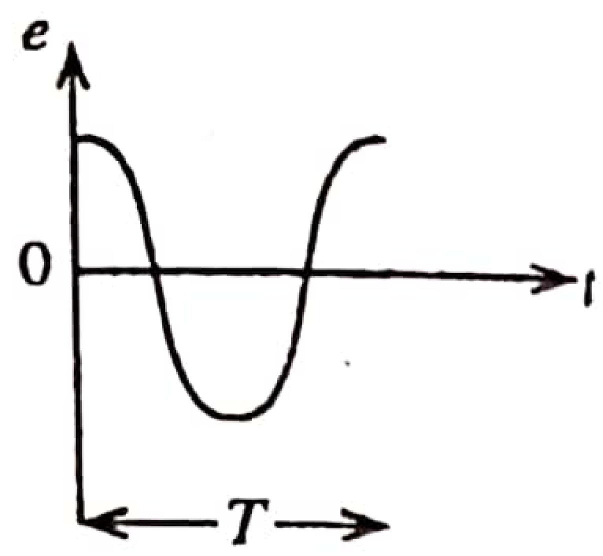
Now the potential of point A is like the previous way. There is no change to the circuit with the cell that has R_3 and R_4. Now how much can be the potential of B? This can get wrong. There is no current across R_1. Therefore, the potential of B should be V_0. As there is no current flow, the potential difference across R_1 should be zero. Then the potential of point B needs to be V_0 and it cannot be zero. If it gets zero, then there will be a potential difference between R_1 (V_0-0). Then there should be a current flow.

Now the potential of A relative to B, $V_A - V_B = R_3 . V_0/(R_3 + R_4) - V_0$

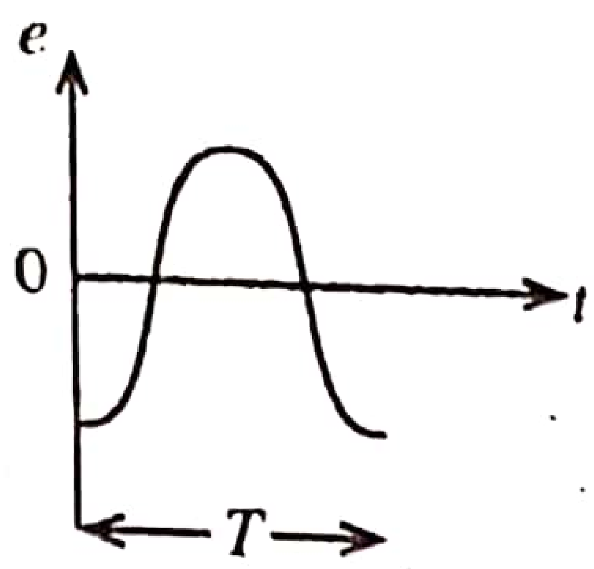
The correct answer is (4). Try without earthing too.

49. A Ferris wheel which consists of two parallel large wooden wheels joined together with metal cross bars as shown in figure, is erected so that the planes of wheels are in the north-south direction, and the cross bars are perpendicular to the direction of the earth's magnetic field B which is horizontal at this location. The Ferris wheel rotates around the horizontal axis passing through the centres of the two wheels at a constant period of rotation T in the direction shown. LM is a metal cross bar which is at the highest position as shown when time $t = 0$. Variation of the induced electromotive force (e) at the end L of the cross bar with respect to the end M with time (t) is best represented by

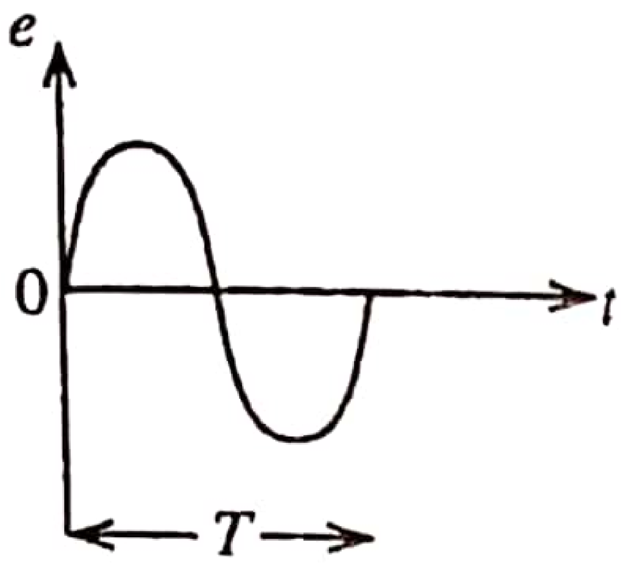




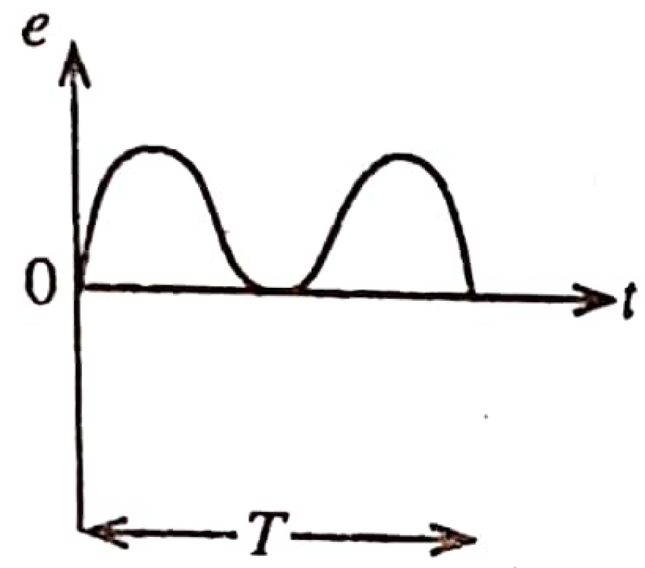
(1)



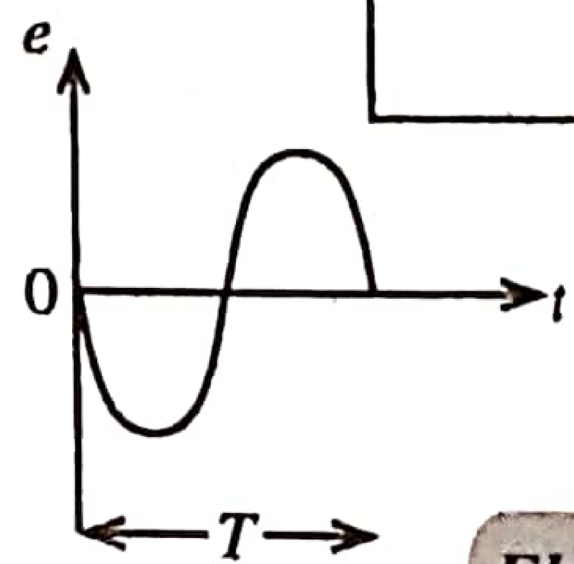
(2)



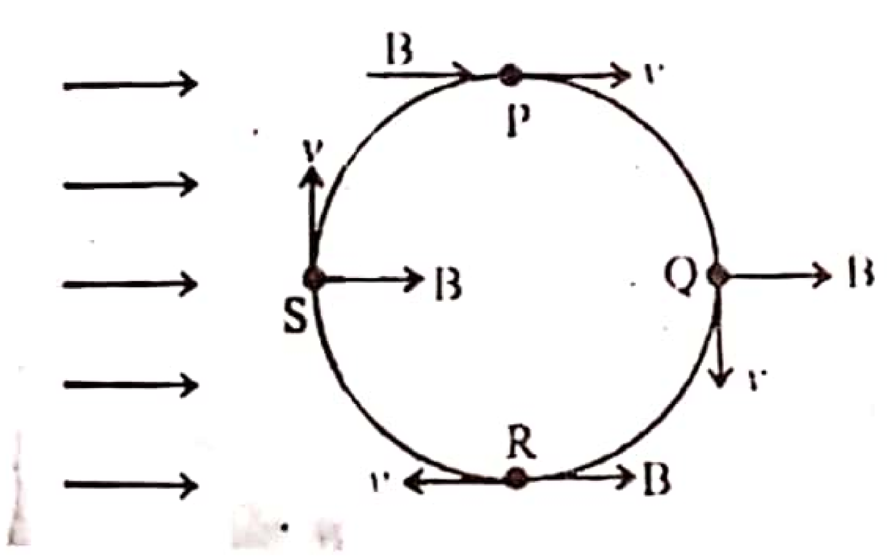
(3)



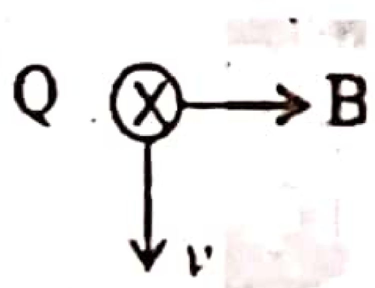
(4)

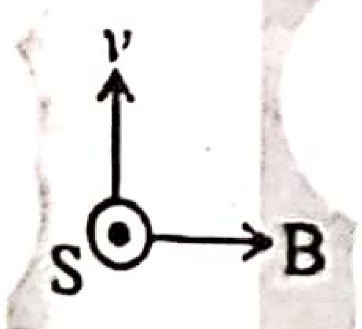


(5)



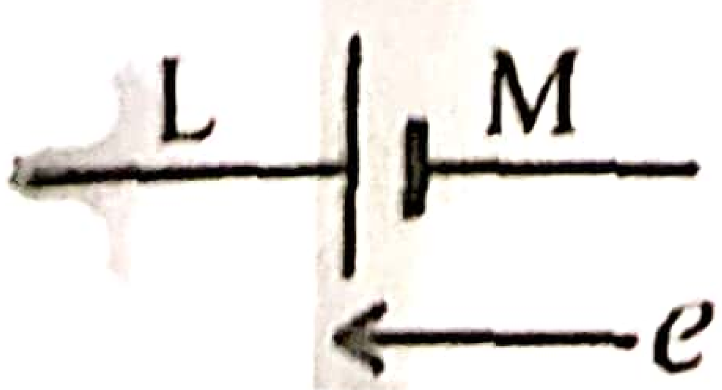
Here a rotating swing has been given to make the question more beautiful. Otherwise, the problem is equivalent to a rotation of a metal rod. Consider a rod which goes in a circle at a uniform horizontal magnetic field. It is enough to consider 4 positions. The rod is perpendicular to the paper.





At the point of P (starting point), v and B are parallel to each other. Then the induced e. m. f is zero. Therefore, the variation should start form zero. From this (1) and (2) can be removed. At point Q, v and B are perpendicular to each other. Therefore, vlB e. m. f in the rod will be directed outside the paper.

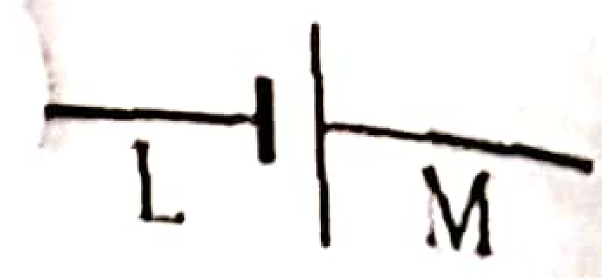
Keep the right thumb perpendicular to the other fingers and rotate the fingers from v towards B. then the thumb will be directed perpendicularly outwards to the paper. As v and B are parallel to each other, vlB force is zero at R. At point S, the direction of v is reversible relative to Q. From that you can remove the variation of (4). Now, vlB force is acting perpendicularly inwards to the paper. Only (3) and (5) will be left out. To find the correct answer out of these two, you need to find whether L end is positive or negative relative to M. Consider the instance of Q. Here the e. m. f is acting along the rod from M to L (out of the paper). Consider the rod as a small cell. If the e. m. f is acting from M to L, then the negative end of the cell should be M whereas the positive end will be L.



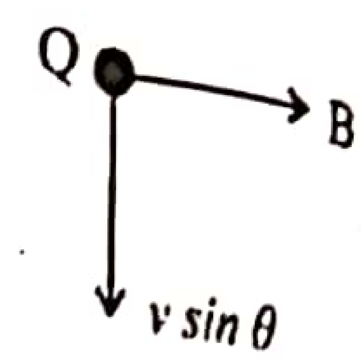
Therefore, L is positive relative to M. So, the correct choice is (3).

You need to start from the positive side. If you think in another way, e. m. f form M to L means if a current is flown, then it should be flown from M to L. So, the charges flow from L to M. Therefore, L is positive relative to M.

You do not even have to look at the instance of S. If you look, then the rod is equivalent to a cell like this.

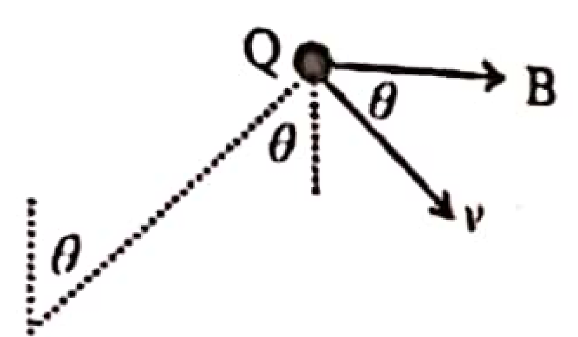


You can remove (1) and (2) as soon as you see. First, the direction of B is drawn perpendicularly to LM. The initial velocity of LM is also given to the direction of B. So, you will get that the correct choice should be either (3) or (5) inherently. When the armature of the A. C dynamo goes one round, then the direction of the induced e. m. f changes. Even if you choose blindly either you need to pick (3) or (5).



The easiest way to decide the direction is to consider the rod as a single cell. To get the directions, use Right Hand Rule. It is simpler than Fleming's rules. You do not have to think of intermediate stages. Always try to grab from the easiest places. If you need to consider an intermediate state, then resolve v. From the horizontal component there is no induced e. m. f. Only from the vertical component an e. m. f is induced.

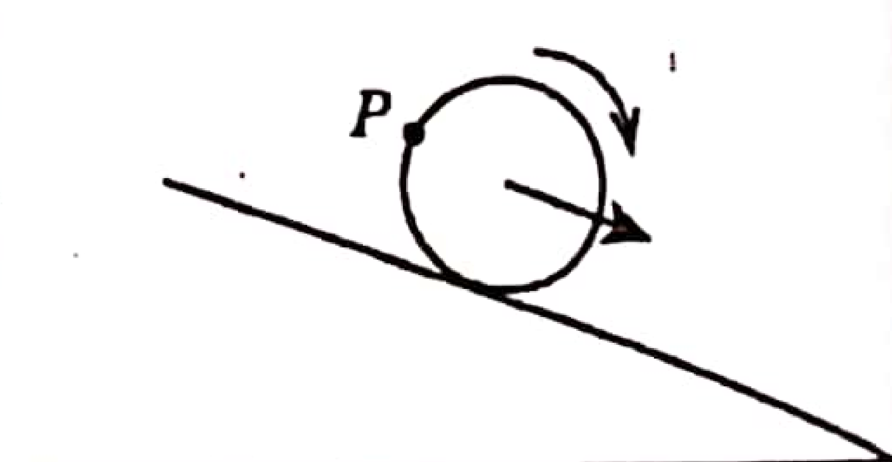
$e = Blv\ sin\theta$



Initially, $\theta = 0$. Then $e = 0$. Even from the above expression you can expect a sinusoidal shape. $\theta = \omega t$.

So, what help is there instead of (3) or (5)?

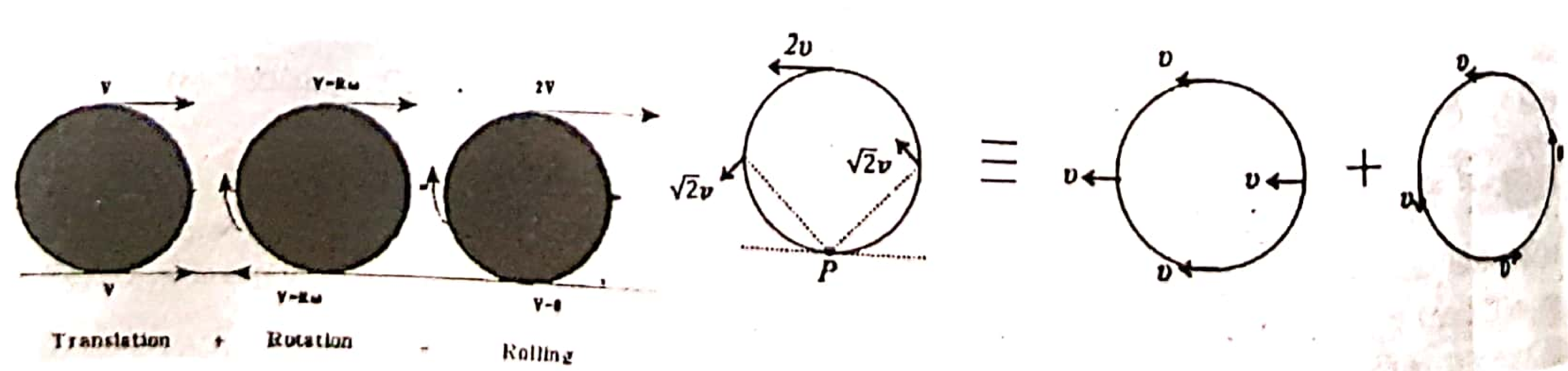
50. A wheel, starting from rest, is allowed to roll down without slipping, along an inclined plane as shown in figure. Which of the following graphs best represents the variation of the magnitude (v) of the velocity of a point P, located on the perimeter of the wheel, relative to the earth with time (t)? (The point P touches the inclined plane at $t = 0$.)

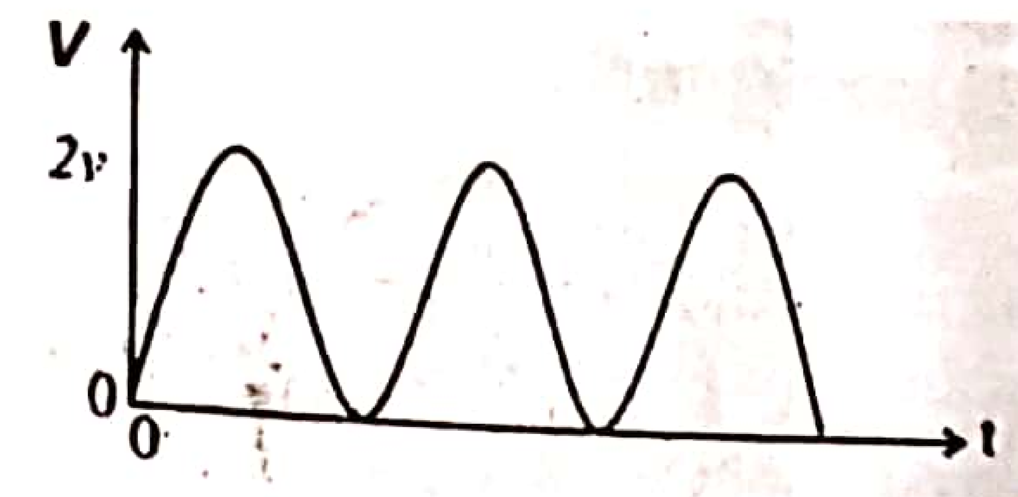


(1) (2) (3) (4) (5)

Rotational Motion

Such a wheel that is rolling on a horizontal floor has been discussed in previous past papers. (Please look at the 18th question of paper 2010 and 57th question of paper 2012-old syllabus)

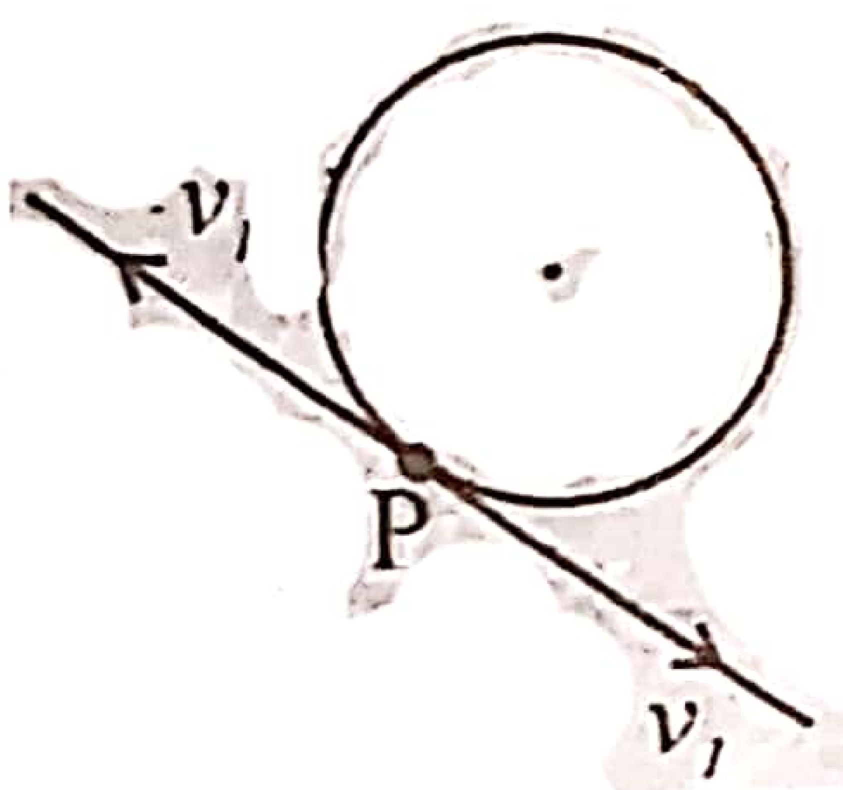




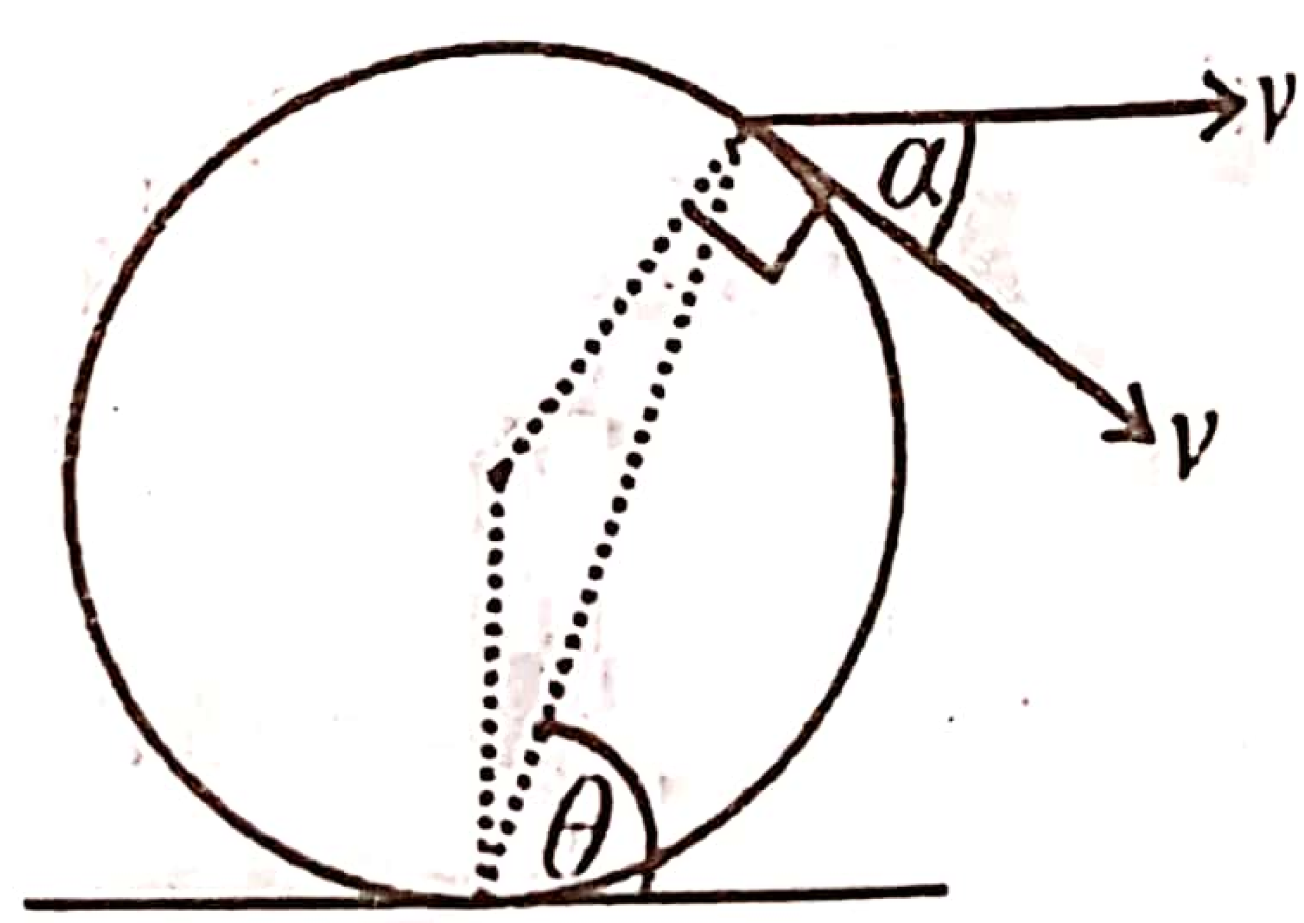
Relative to the ground, the magnitude of the velocity of point P varies from zero to 2v. When it is touching the ground, the velocity of point P is zero. When it goes to the top of its path, then the velocity gets 2v. Every other magnitude of the velocity is lesser than 2v. Therefore, the velocity of point P relative to the earth (V) with time (t) of a wheel that rolls on a horizontal floor varies like this. The maximum magnitude (velocity amplitude) does not change. It is 2v.

Now let us consider the wheel in an inclined plane. Now as the wheel is accelerating downwards, v gets increased with time. Then on each circle, the value of 2v also gets increased. That means the velocity amplitude is not a constant. It increases with the time. The other thing that happens here is that, as the wheel accelerates, the time taken for the wheel to go one round gets decreased gradually. Quickly P goes one round.

The gradual increment of the velocity amplitude and the gradual reduction of time taken for a round is represented by the graph of (5). Even in graph (2), the velocity amplitude is increased, the time taken for one each round is equal. (3) is also like (2). The only difference is that the initial velocity of point P is not zero relative to the earth. In (4), in each round of time frames, the time is gradually increased.



Even it rolled in an inclined plane, when point P touches the plane, its velocity is zero relative to the earth. At each instance, the velocity at the centre of mass of the wheel and rotating speed of the mid-point of the wheel across a perpendicular axis is equal and opposite.



If you think in another way, as the wheel is not sliding, when the point P touches the inclined plane, it is at instant rest relative to the earth. If needed, an expression for the magnitude of the velocity of P point on a horizontal floor can be obtained.

The angle between the two velocity vectors (α) can be shown as $180^0 - 2\theta$. Try out and see. Now we will get the resultant of the two velocity vectors by using $R^2 = P^2 + Q^2 + 2PQ \cos \theta$

$V^2 = v^2 + v^2 + 2v^2 \cos (180^0 - 2\theta) = 2v^2 [1+ \cos (180^0 - 2\theta)]$

When $\theta = 0^0$, then $\cos (180^0 - 2\theta) = \cos (180^0) = -1$; $V = 0$

When $\theta = 90^0$, then $\cos (180^0 - 2\theta) = \cos (0^0) = +1$; $V^2 = 4v^2$; $V = 2v$

When $\theta = 45^0$, then $\cos(180^0 - 2\theta) = \cos(90^0) = 0$; $V^2 = 2v^2$; $V = \sqrt{2}v$

When the above expression is plotted with θ (ωt), you will get the shape of the above.

When the wheel is rolled in the inclined plane also you can use the above expression. The only difference is v is increased with the time. If the wheel comes down with an acceleration of a, then after the initial rest, let v be the velocity of the centre of mass after a time t. Then $v = at$ and

$V^2 = 2a^2 t^2[1+ \cos(180^0 - 2\theta)]$; $V = [1+ \cos(180^0 - 2\theta)]^{1/2}$

In this equation there are two variants as t and θ. V is increased with t. It is clear. Always θ varies in a circle. It varies repetitively from $\theta = 0^0$ to 180^0. From this you get the bubbly parts. But the height (amplitude) of a bubbly part is increased with the time and when the wheel comes down, quickly these rounds are gone (as the speed increases). The time taken for a bubbly part also gradually decreased. It is clear from the above expression that for whatever the value of t, if $\theta = 0^0$ (or 180^0), then $V = 0$.