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The Most Complete Guide to Stable Diffusion Parameters

Unlock the full potential of Stable Diffusion with these parameters

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Negative Prompt

A negative prompt is exactly what it sounds like – it's the **opposite of a prompt**. You input is what you **DO NOT want Stable Diffusion to generate**

This is a **very powerful but underused** feature of Stable Diffusion, and it can assist you in achieving results that would take way more time to reach by just tweaking the positive prompt.

Prompt: **Photo of a burger**

Same prompt with
general negative prompts

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General Negative Prompts

lowres, error, cropped, worst quality, low quality, jpeg artifacts, out of frame, watermark, signature

Negative prompts for **people portraits**

deformed, ugly, mutilated, disfigured, text, extra limbs, face cut, head cut, extra fingers, extra arms, poorly drawn face, mutation, bad proportions, cropped head, malformed limbs, mutated hands, fused fingers, long neck

Prompt: **Photo of a woman smiling**

No negative prompts

With people potrait negative prompts

Negative prompts for photorealistic images:

illustration, painting, drawing, art, sketch

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Prompt: **Elephant wearing a party hat**

No negative prompts

Photorealistic negative prompt

Steps

Stable Diffusion creates an image by starting with a canvas full of noise and denoise it gradually to reach the final output. This parameter controls the number of these denoising steps. **Usually, higher is better but to a certain degree. The default we use is 25 steps which should be enough for generating any kind of image.**

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Here’s a general guide on what step number to use for different cases:

- If you’re testing a new prompt and want to have fast results to tweak your input, **use 10-15 steps**
- When you find the prompt you like, **increase the steps to 25**.
- In case you’re creating a face or an animal with fur or any subject that has detailed texture, and you feel the generated images are missing some of these details, **try to bump it up to 40!**

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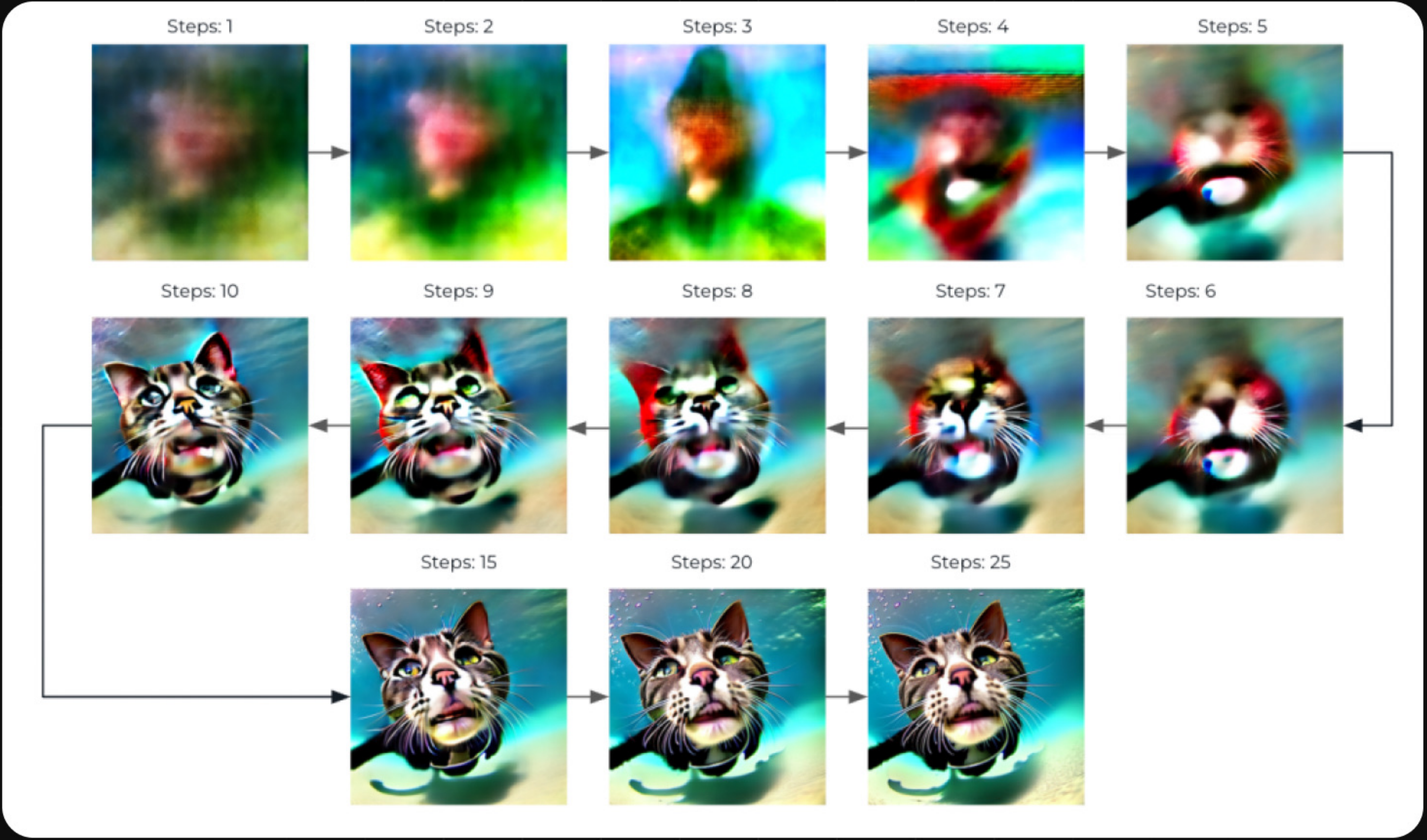

Steps: 1
Steps: 2
Steps: 3
Steps: 4
Steps: 5
Steps: 10
Steps: 9
Steps: 8
Steps: 7
Steps: 6
Steps: 15
Steps: 20
Steps: 25

Samplers

Diffusion models work by **denoising a starting noise canvas**. This is where Diffusion samplers come to work. In simple terms, these **samplers are algorithms that take the generated image after each step** and compare it to what the text prompt requested, and then **add a few changes to the noise till it gradually reaches an image that matches the text description**.

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the three most used samplers

- Euler A
- DDIM
- DPM Solver++

You can try the three and see what fits your prompt better since **there is no rule on what sampler to use**, but these three are very fast and capable of producing coherent results in 15-25 steps

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Euler a

DPM Solver++

There is **only one noticeable difference** worth mentioning, see how Euler a results – compared to DPM Solver++– **have smoother colors with less defined edges**, giving it more of a “dreamy” look.

CFG guidance scale

This parameter can be seen as the **"Creativity vs. Prompt"** scale. Lower numbers give the AI more freedom to be creative, while higher numbers force it to stick more to the prompt.

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Going lower than 5 is generally not recommended as the images might start to look more like AI hallucinations, and **going above 16 might start to give images with ugly artifacts**

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So when to use **different CFG scale values?**

- **CFG 2 – 6:** Creative, but might be too distorted and not follow the prompt. Can be fun and useful for short prompts
- **CFG 7 – 10:** Recommended for most prompts. Good balance between creativity and guided generation
- **CFG 10 – 15:** When you're sure that your prompt is detailed and very clear on what you want the image to look like
- **CFG 16 – 20:** Not generally recommended unless the prompt is well-detailed. Might affect coherence and quality
- **CFG >20:** almost never usable

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Seed

The seed is a number that decided **the initial random noise** we talked about previously, and since the random noise is what determines the final image, i**t is the reason you get a different image** each time you run the exact same prompt on StableDiffusion, and why you get **the same generated image if you run the same seed with the same prompt multiple times.**

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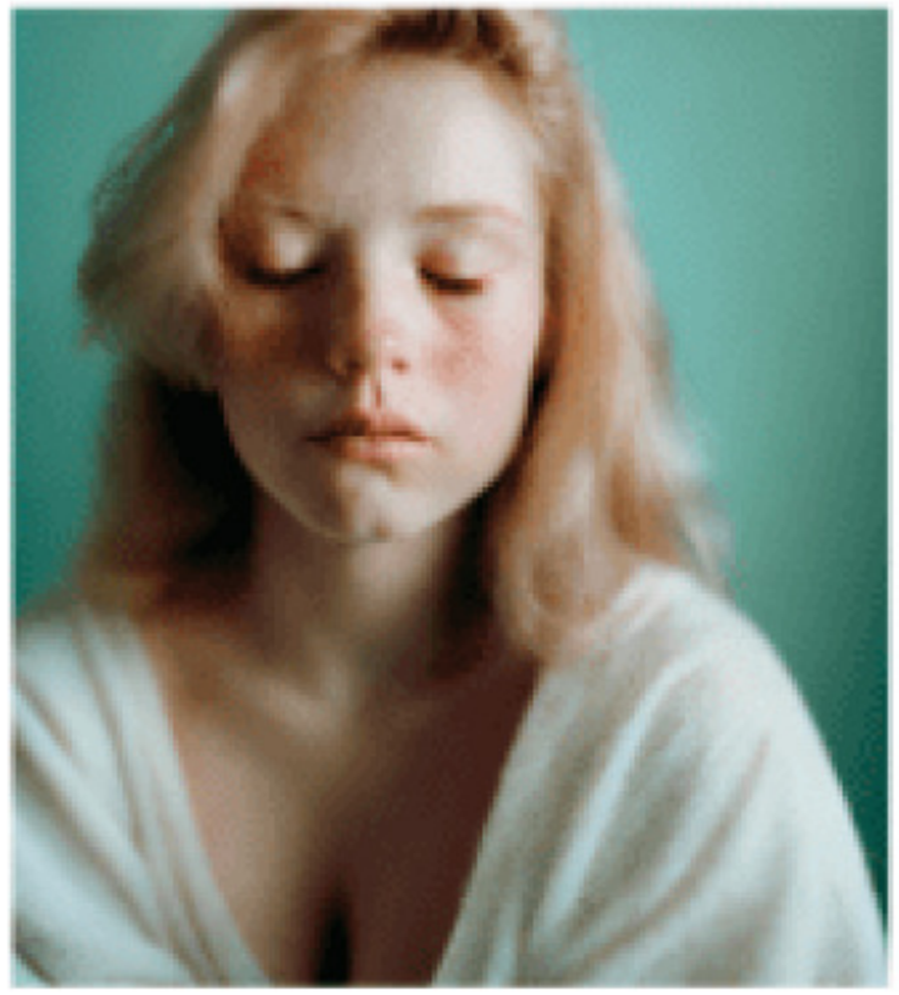
Since the same seed and prompt combo gives the same image each time, we can use this property to our advantage in multiple ways:

portrait of a beautiful woman with blonde hair and blue eyes

portrait of a beautiful woman with blonde hair and blue eyes, smiling

portrait of a beautiful woman with blonde hair and blue eyes, crying

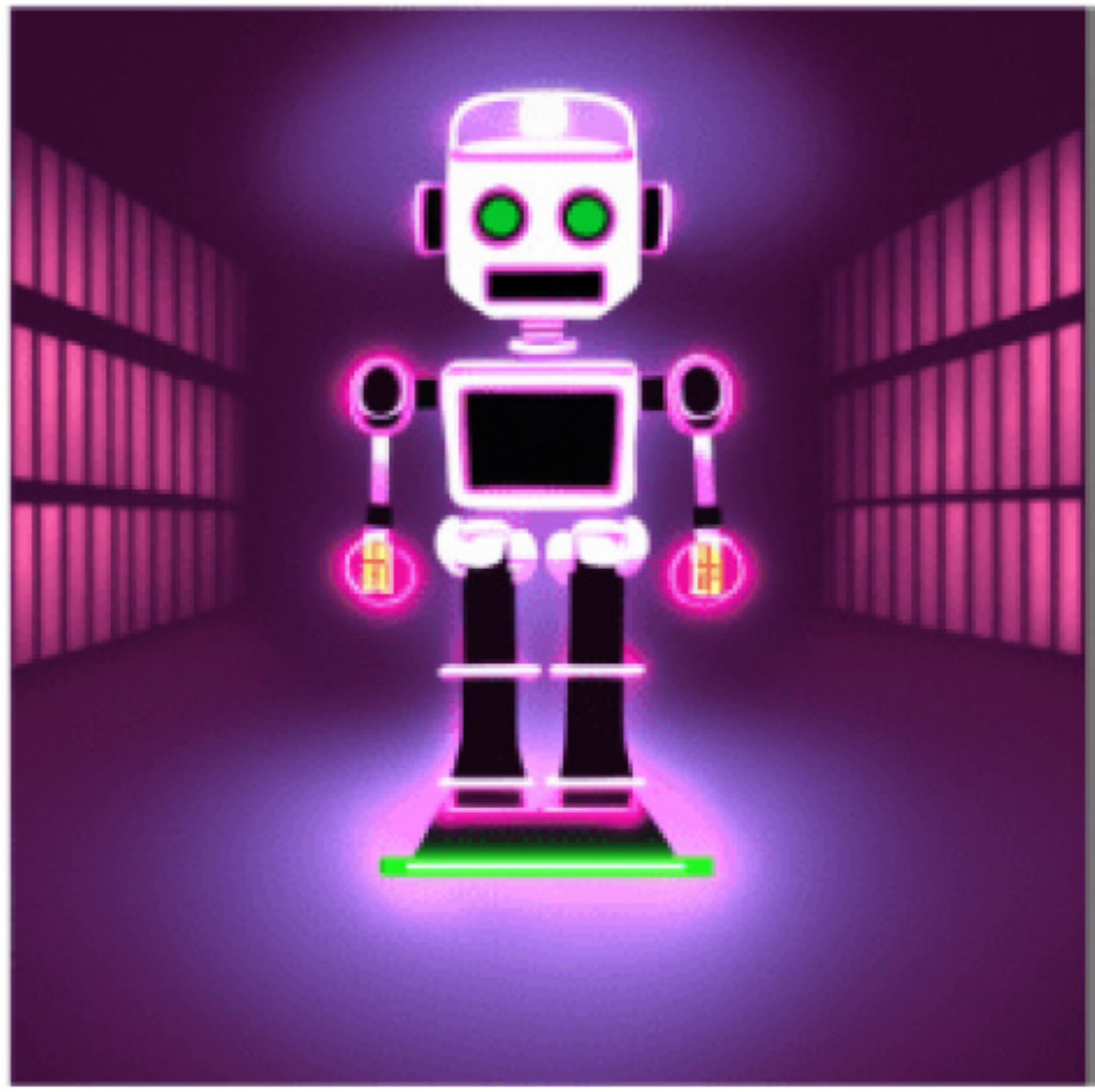
portrait of a beautiful woman with blonde hair and blue eyes, eyes closed



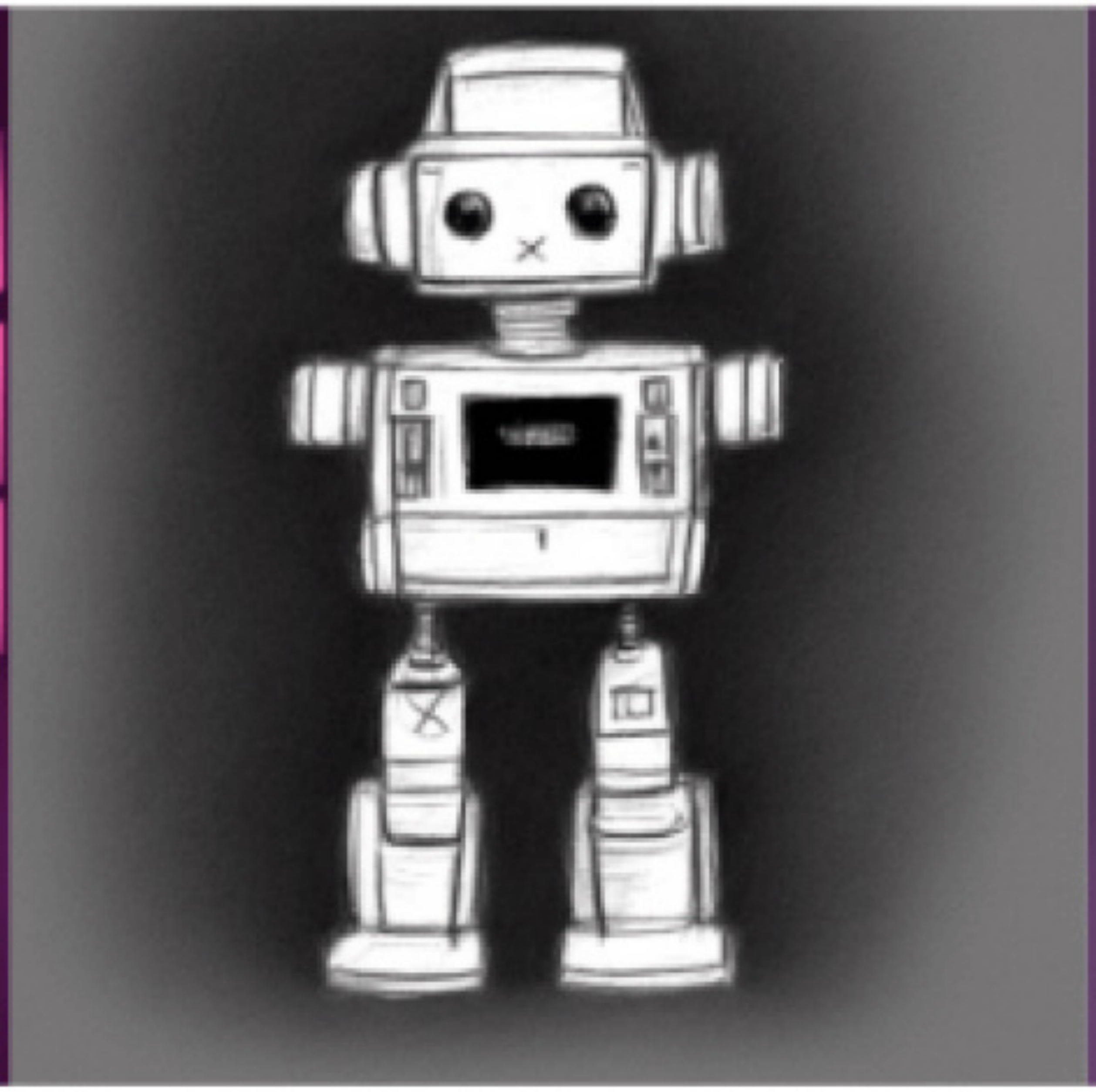
Control specific features of a character:

in this example, we changed the emotion, but this can also work for other physical features like hair color or skin color, but the smaller the change the more likely it will work

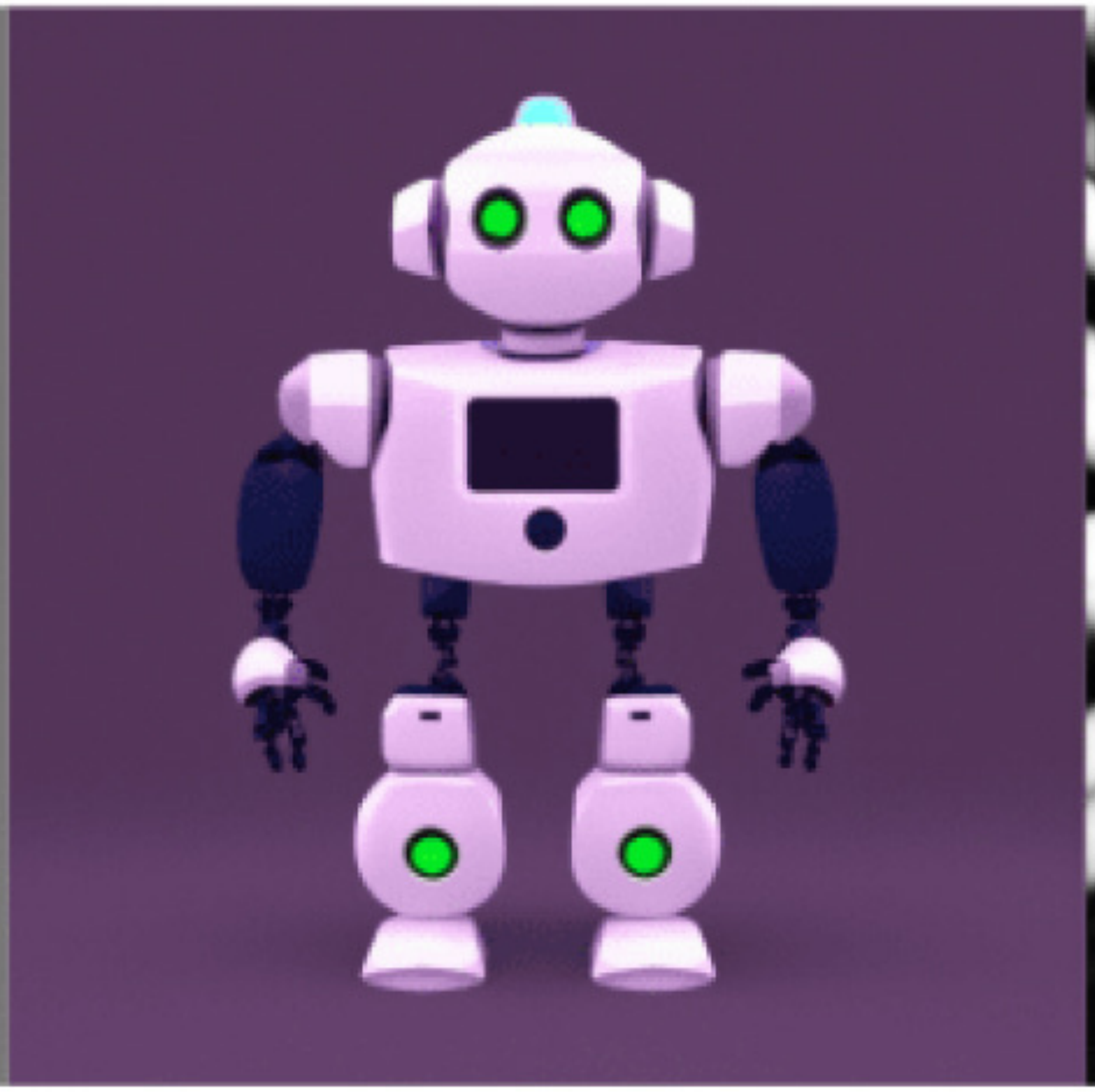
A cute robot, standing in a dark room, **synthwave**



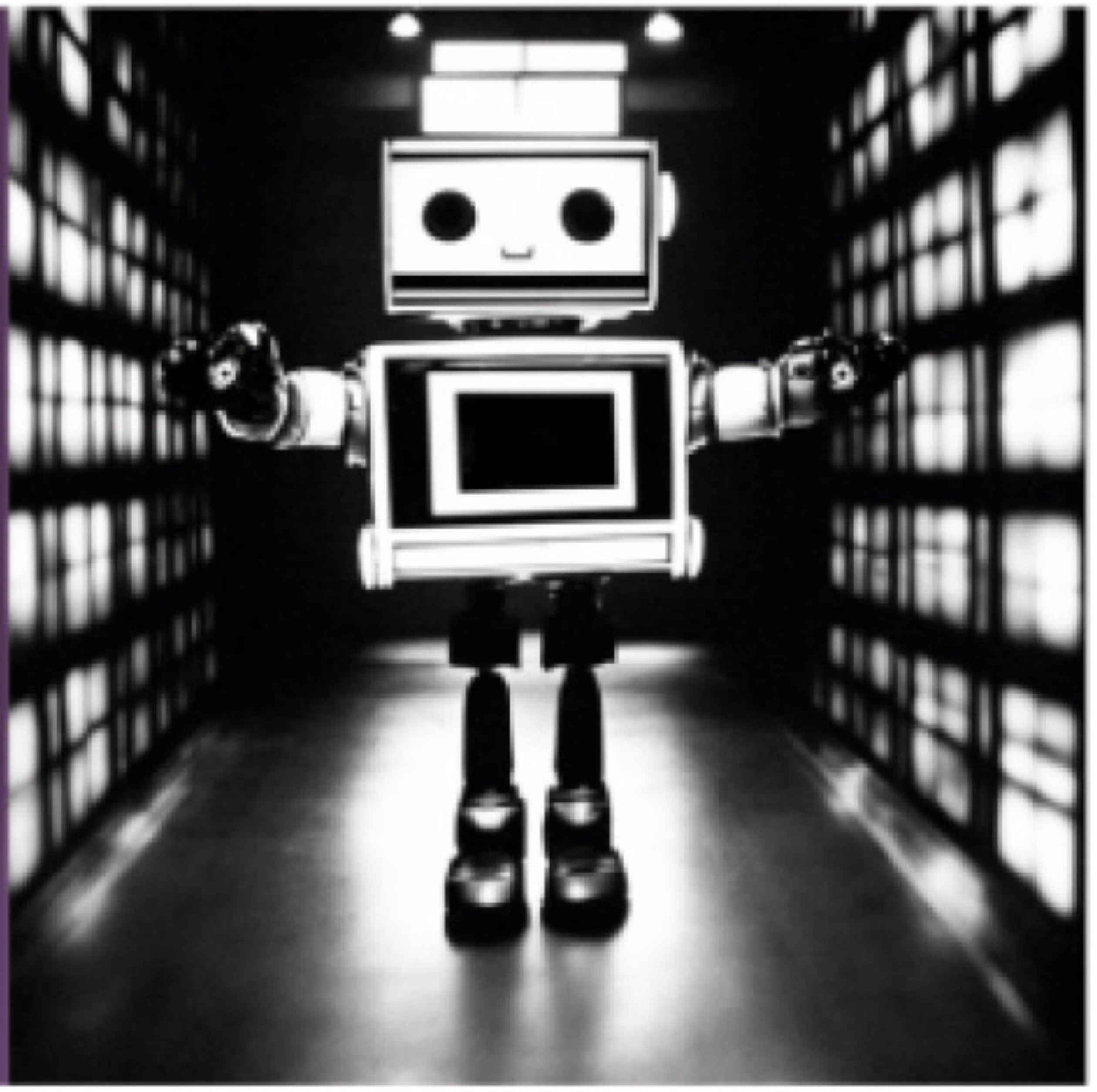
A cute robot, standing in a dark room, **pencil sketch**



A cute robot, standing in a dark room, **low poly**



A cute robot, standing in a dark room, **film noir**



Testing the effect of specific words:

If you wonder what a specific word is changing in the prompt, you can use the same seed with a modified prompt to test it out, it’s good practice to test prompts this way by changing a single word or phrase each time

By Van Gogh

By Frida Kahlo

Pixar render

By Greg Rutkowski

Change style:

If you like the composition of an image, but wonder how it would look in a different style. this can be used for portraits, landscapes, or any scene you create.

Img2img Parameters

The Img2img feature works the exact same way as txt2img, the only difference is that **you provide an image to be used as a starting point instead of the noise generated by the seed number.**

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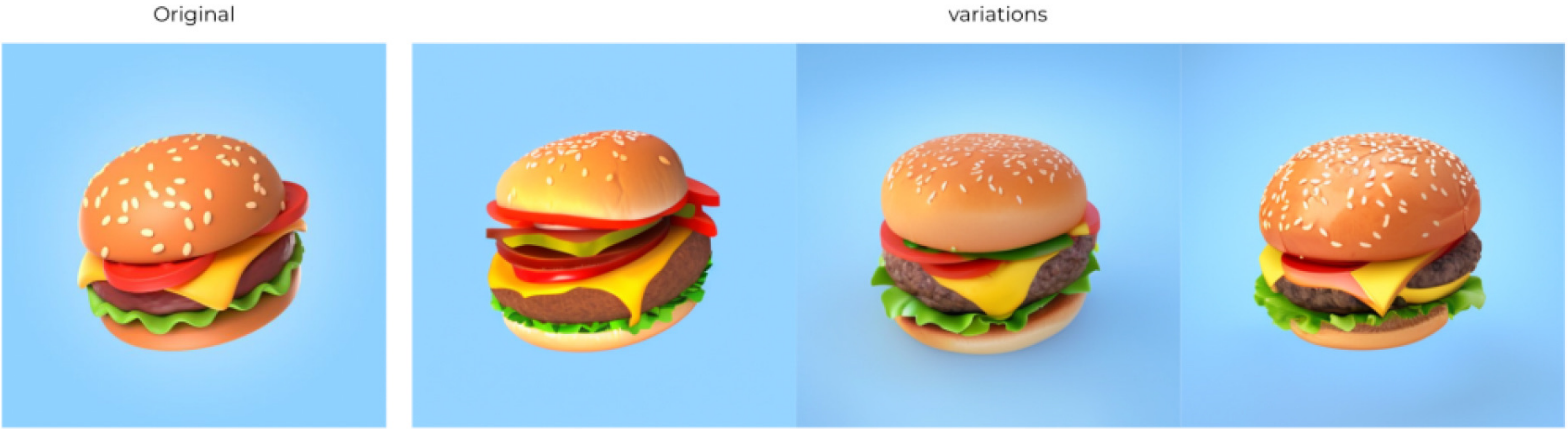
Noise is added to the image you use as an init image for img2img, and then the diffusion process continues according to the prompt.

The amount of noise added depends on the "Strength of img2img" parameter, which ranges from 0 to 1, where 0 adds no noise at all and you will get the exact image you added, and 1 completely replaces the image with noise and almost acts as if you used normal txt2img instead of img2img.

So how to decide what strength to use?

Here is a simple guide with examples



To create variations of an image, the suggested strength to use would be **0.5-0.75** and with the same prompt.

Init image

Ran img2img once with strength of 0.7

Ran img2img 4 times with looping at strength of 0.25

To change an image style while keeping it similar to the original, **you can use a lower-strength img2img multiple times,** and get way better image fidelity compared to a single img2img with higher strength.

For this example we used a strength of **0.25 for 4 times**, so each time we generate the image we re-insert the generated image into the img2img and rerun it with the same prompt and strength till we get the style we need.


That's all Folks!