



STABLE DIFFUSION PROMPT BOOK

**Innovating Marketing Practices
through AI Image Generation**

by Georg Neumann, Veronika Hackl and Nabil Khaled



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01 | INTRODUCTION

Introduction

Get ready for a unique introduction to an amazing journey. I started this book from scratch, erasing all of Stable Diffusion from my device. Why? To guide you, step-by-step, from the exciting start of installing AUTOMATIC1111 on your device to the discovery of extra plugins, styles, embeddings Loras, and beyond.

This book will take you from knowing nothing about Stable Diffusion to creating stunning visuals. Whether you're a designer, an artist, or someone in marketing or advertising, this book is a wellspring of inspiration. I've explored every aspect of this field - from the joys of creation to the details of analytics.

This book is for the brave among you ready to explore the exciting world of AI and image-generation. For those who've walked my path, this journey will change your life forever. You'll come out of it, not just informed, but completely transformed.

And there's even more. The images in this book aren't just decoration, they're part of the story. Each was carefully created during the writing process to both teach you and captivate your imagination. They're more than illustrations, they're unique guides on this journey, created to give you an experience like no other. So, get ready to start a journey that will educate and delight you like never before.



Get to Know the Experts Guiding Your AI Journey



Georg Neumann is an expert in integrating artificial intelligence (AI), particularly image AI and Stable Diffusion, into design and marketing. With a decade of industry experience, he co-runs a successful design and branding agency and provides specialized AI training to other professionals.

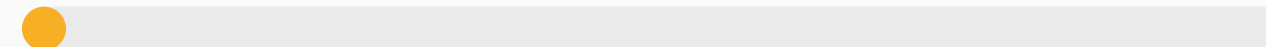


Veronika Hackl, an entrepreneur from Passau, leverages her international experience and a strong belief in cultural diversity. She's currently pursuing a doctorate focusing on AI in higher education and co-hosts an AI Marketing Bootcamp alongside Georg Neumann.



Nabil Khaled is a marketer and brand growth consultant with a keen interest in AI trends. As the founder and director of a digital marketing agency, he continuously explores how AI can influence and shape modern marketing strategies.

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What are AI Image Generation Models?

An AI image generation model, capable of creating novel images, learns from existing image datasets. Using specific text inputs, these models can produce unique visuals. This book focuses on Stable Diffusion, a type of image-generation model based on Generative Adversarial Networks (GANs) with a unique twist.





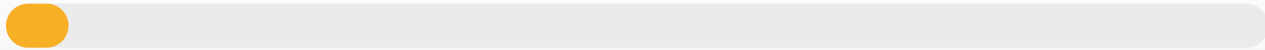
Generated on Midjourney

Other Popular Image Models

DALL-E, by OpenAI, generates images from text prompts, known for creating surreal visuals by combining unconventional elements. Midjourney, by Midjourney Research Lab, emphasizes aesthetics and emotional expression, offering customization of image size and detail levels through AI steps. Each model has unique features, strengths, and limitations, and the choice depends on user requirements. Both OpenAI and Midjourney Research Lab continuously refine their models to enhance image generation capabilities.

A Comparison between image generation models

Model	Creator/Company	Description	Features	Style	Speed	Availability
DALL-E	OpenAI	A text-to-image model that generates high-quality images based on textual prompts	<ul style="list-style-type: none">- Advanced generative capabilities- Large-scale training on diverse images- Supports various image resolutions	Moderate	Moderate	Available from OpenAI
Midjourney	Midjourney Research Lab	A text-to-image model known for its expressive and artistic style in generating images	<ul style="list-style-type: none">- Emphasizes aesthetic and artistic aspects- Produces expressive and moody images	High	Moderate	Available from Midjourney
Stable Diffusion	Stability AI	A text-to-image model that focuses on stability and robustness in generating detailed images	<ul style="list-style-type: none">- Offers customization options for image generation- Provides in-depth image customization- Supports detailed image conditioning	Moderate	Moderate	Available from Stability AI



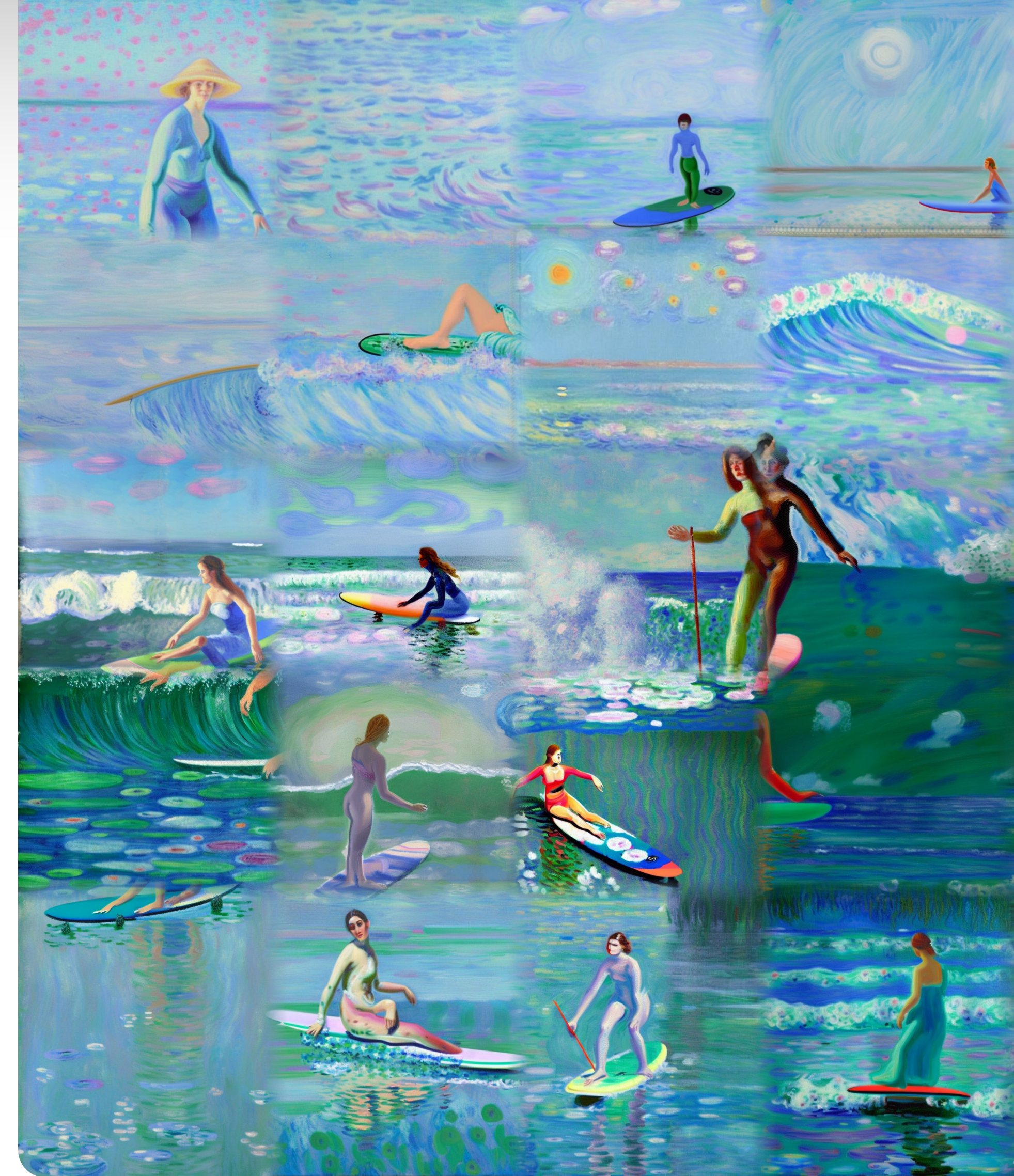
DreamBooth

Dreambooth is a groundbreaking tool that enables the fine-tuning of already pre-trained AI models. Particularly for Stable Diffusion, Dreambooth opens up new possibilities by allowing the model to be further developed with one's own image material. This creates the prerequisite for implementing personal styles or training with specific faces and objects.

In comparison to competitors such as Midjourney or DALL-E, Stable Diffusion gains a significant advantage through the integration of Dreambooth. While there's only one version of the others, Stable Diffusion offers the choice of thousands of models or even the option to train completely own models.

The open-source nature of these technologies promotes exchange within the community and enables continuous development and improvement. However, ethical concerns and the potential for misuse must also be taken into account.

Please note that we do not delve deeper into the training of image models with Dreambooth in this prompt book. Those with further interest can book a Deep Dive with Georg Neumann in the AI Marketing Bootcamp.





Installation

Starting with Stable Diffusion, you have the option to use web-based platforms, Python code, or the dedicated GUI. This book will concentrate on the GUI, which involves installing it on your computer or running it via a cloudservice such as rundiffusion.com. Our detailed guide on page 188 will assist you, if you want to install it locally.

It's worth noting that I've enjoyed various web-based apps during my journey, all of which performed well. However, running Stable Diffusion locally on your PC or Mac retains the joy, liberty, and absence of censorship that web-based apps might limit.

02

UNDERSTANDING THE BASICS



Getting started

The first variables you should establish are the resolution: In the simplest terms, if you have a powerful GPU, go for 768 x 768 if your GPU is a more towards average power, you can just stick to 512 x 512.

Secondly; As a beginner I would change the sampling method to **DDIM**, it's usually on **Euler a** by default. Don't worry, you'll be learning more about this later!

Start getting familiar with the Graphical user interface for Stable Diffusion, this environment is where you will create all the magic. Try to check out all the variables, but don't get too excited just yet.

Prompting

Let's jump right into it: A prompt, in image generation, is the creative spark guiding a machine learning model to produce an image. Think of it as ordering a meal in a restaurant: "I'd like a vivid sunset landscape, complete with serene beach, palm trees, and a couple strolling hand-in-hand". Just like ordering, a clear, well-defined prompt helps cook up the perfect AI-generated image, while a vague one might give you something you didn't quite expect. Be the chef of your AI creativity!



Examples Of prompts:

Beginner



Prompt: Aerial shot of a city, skyscrapers, sea, sunny day, professional photo, intricate details

Intermediate



Prompt: (a design of a Coffee logo:1.3), featuring a mushroom cloud coming out of a cup, (the cloud looks like brains:1.1), (art by mcbess:1.1), full colour print, vintage colours, 1960s, Vector art, Vivid colors, Clean lines, Sharp edges, Minimalist, Precise geometry, Simplistic, Smooth curves, Bold outlines, Crisp shapes, Flat colors, Illustration art piece, High contrast shadows, Technical illustration, Graphic design, Vector graphics, High contrast, Precision artwork, Linear compositions, Scalable artwork, Digital art

Advanced



Prompt: A photo of jeff bezos, (a mysterious magician:1.3), (soft lighting:1.2), (holding lots of money in his hands:1.4), (with a colorful lighting effect around him:1.3), (detailed smart:clothing:1.4), (superrealistic:1.3), cannon r50,5mm lens, f1.8, depth of field, (natural_landscape_background:1.6), soft lighting, detail, hdr, symmetrical face, perfect facial features, realistic_money, detailed facial hairs, (realistic bokeh lighting:1.4), (perfect hands:1.5), (professional photography:1.3)

Prompting For Beginners

Consider each prompt as a person: the first part is the brain, driving the entire operation, the second part is the heart, while the rest act as limbs and arms, forming a complete entity. Importantly, the 'brain' or first part of the prompt has the most impact on the outcome. If the prompt is lengthy, words further along carry less influence.



Break it down

Let's break it down into three basic components: **brain, heart, and body**. A popular format often suggested is “A [type of picture] of a [main subject], [style/context cues]”. This blueprint offers clear context, guiding the model to capture your envisioned image concept accurately.

E.g. A painting of a woman wearing a red coat in the wilderness.



Steps: 30, Sampler: Euler a, CFG scale: 7, Seed: 2077891381, Size: 768x768,
Model hash: 4711ff4dd2, Model: v-768_1-2nonema-pruned

Break it down

You can take that a step further by being more descriptive in each of the previously mentioned sections.

Example:

“A drawing, of a cute kid, playing with a ball, pencil art, masterpiece”

Generally, the more clear and well-worded your additional descriptive keywords are, the better the result.



Steps: 30, Sampler: Euler a, CFG scale: 7, Seed: 3630866767, Size: 768x768,
Model hash: 4711ff4dd2, Model: v-768_1-2nonema-pruned



Technically speaking

Prompting in AI image-generation is similar to issuing a command in programming, with a specific syntax that the model understands. Like nested function calls, the prompt follows a hierarchical structure, where the initial segment is crucial. Despite upgrades like V2.1, the AI's interpretation is limited by its understanding of the syntax and semantics, making single-keyword prompts less effective. The art of prompt creation is akin to writing clear, succinct code, balancing high-level directives with fine-tuned parameters.

In Simpler terms

Treating the image model as a human interpreter can lead to disappointing results. Despite significant improvements from V1.5 to V2.1, single keyword prompts like “Dog” typically yield unsatisfactory outcomes. However, the upside in V2.1 is that even basic prompts can produce reasonably good results.

Look at this comparison chart using V2.1
Remember: In V1.5 this would be much worse.

The top images, the prompt is “Dog”
The bottom images, the prompt is “RAW candid cinema, dog, ((remarkable color)), (ultra realistic)” taken from **Kamph’s example**.

You can clearly see that by using one very short prompt with the right wording, order and punctuation, you can have remarkable results.

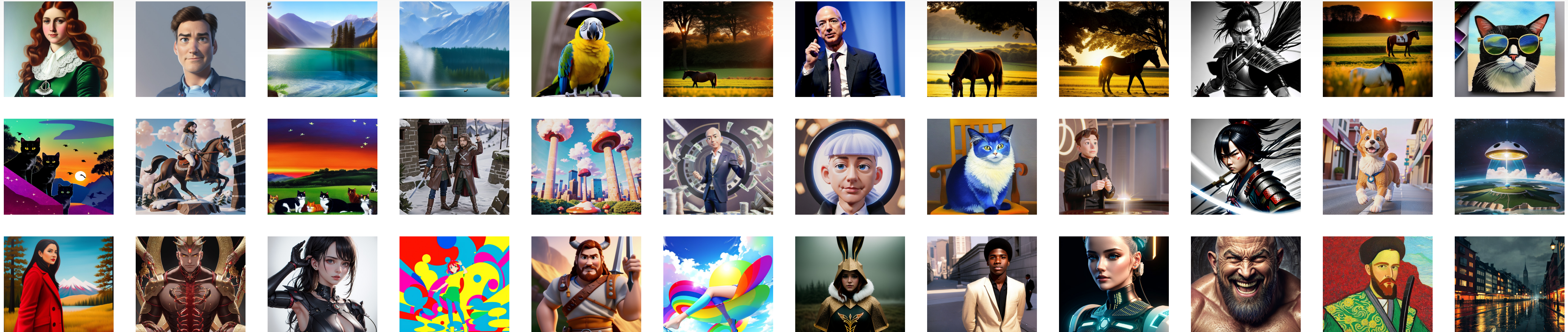


Let's take a step back and delve deeper into the anatomy of the prompt.

First, begin by asking yourself these questions and take note of the highlighted keywords:

- What's the overall **medium**: realistic photo, painting, or digital artwork?
- If an **artistic style**, does it resemble a 2D illustration, 3D render, or album cover?
- If a photo, what **type**: portrait, closeup, specific camera/lens features?
- What's the **central subject/element**: human animal or object
- What's the **setting**: rooftop, outdoors, bedroom or underwater?
- How's the **lighting**: bright, soft, bokeh, backlight, dark or even a popular artificial light source?
- Do you want it to be inspired by a specific **artist or reference**?





Medium

First “The Brain”: This refers to the medium or category of the artwork, including examples such as portrait, illustration, concept art, underwater, digital painting, sketch, drawing, and 3D render.

Subject

This aspect is straightforward. Your subject can encompass a wide range of possibilities, including men, women, children, animals, toys, figurines, cars, and any other subject that can be linguistically defined.

Style

This offers a deeper exploration and refinement of the original medium. Examples of styles include Hyperrealistic, Pop-art, Modernist, Realistic, Minimalistic, Photorealistic, Pointillist, Fauvist, Symbolist, and Baroque.

Setting

The scene setting can be classified into two types: descriptive (sand dunes, rooftops, underwater, clouds) and specific (Spain, NY cityscapes, crowded streets of Cairo, Pixar studio, etc.).

Quality/Resolution

This variable greatly affects image details. Examples: Maya Render, Octane Render, Unreal Engine, 8k, 4k, Vray, sharpness, focus, and depth of field.

Lighting & Color

Subject perception depends on lighting and color. Learn to use different lighting types, like soft, sunlight, studio, bokeh, and colored, and color schemes like vintage or monochrome for impactful imaging.

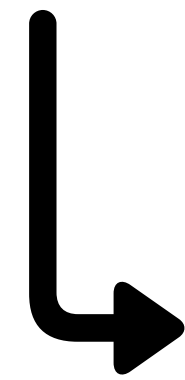
Artist

This is where the excitement truly begins. You can draw inspiration from any notable artist of your preference, be it acclaimed photographers like Peter McKinnon, visual artists, or distinguished painters such as Van Gogh.

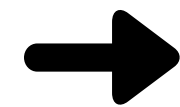
Website/reference

Going a step further, you can draw inspiration from popular websites known for their distinct styles or standards, including Deviant Art, ArtStation, and Pixabay.

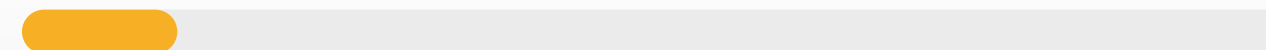
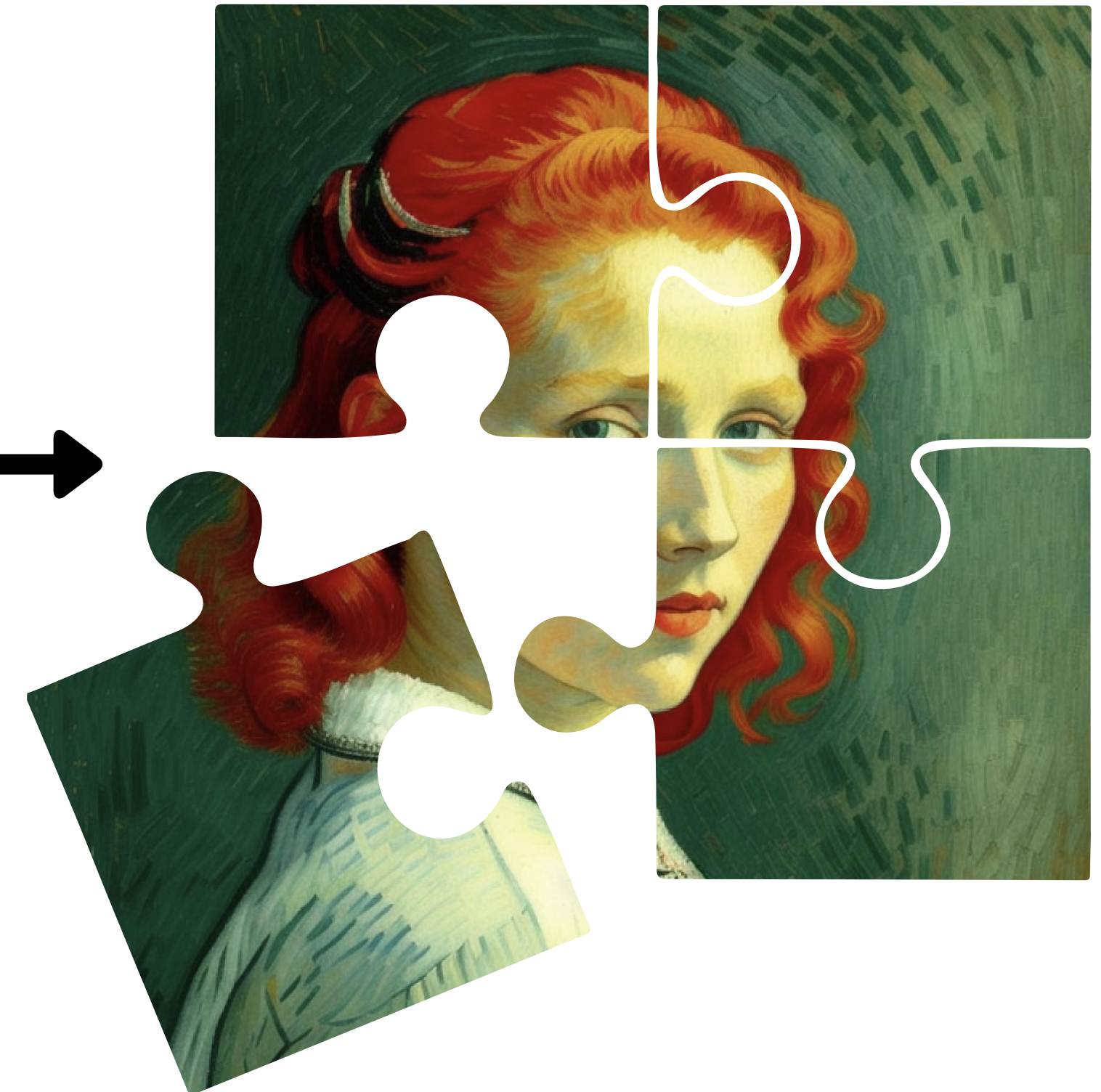
Puzzle It Together



In the most basic form this
is one way to look at it:
medium, **subject**,
descriptors, **artist**



A portrait of a young lady,
with red hair, masterpiece
painting, by Van Gogh



Step it up a Notch



Medium (State) Subject Setting
Color Style/descriptor Website
reference +Descriptor Visual quality

A digital illustration of a happy duck,
plain background, warm sunlight,
modern 2D digital illustration, deviant
art, masterpiece, 8k



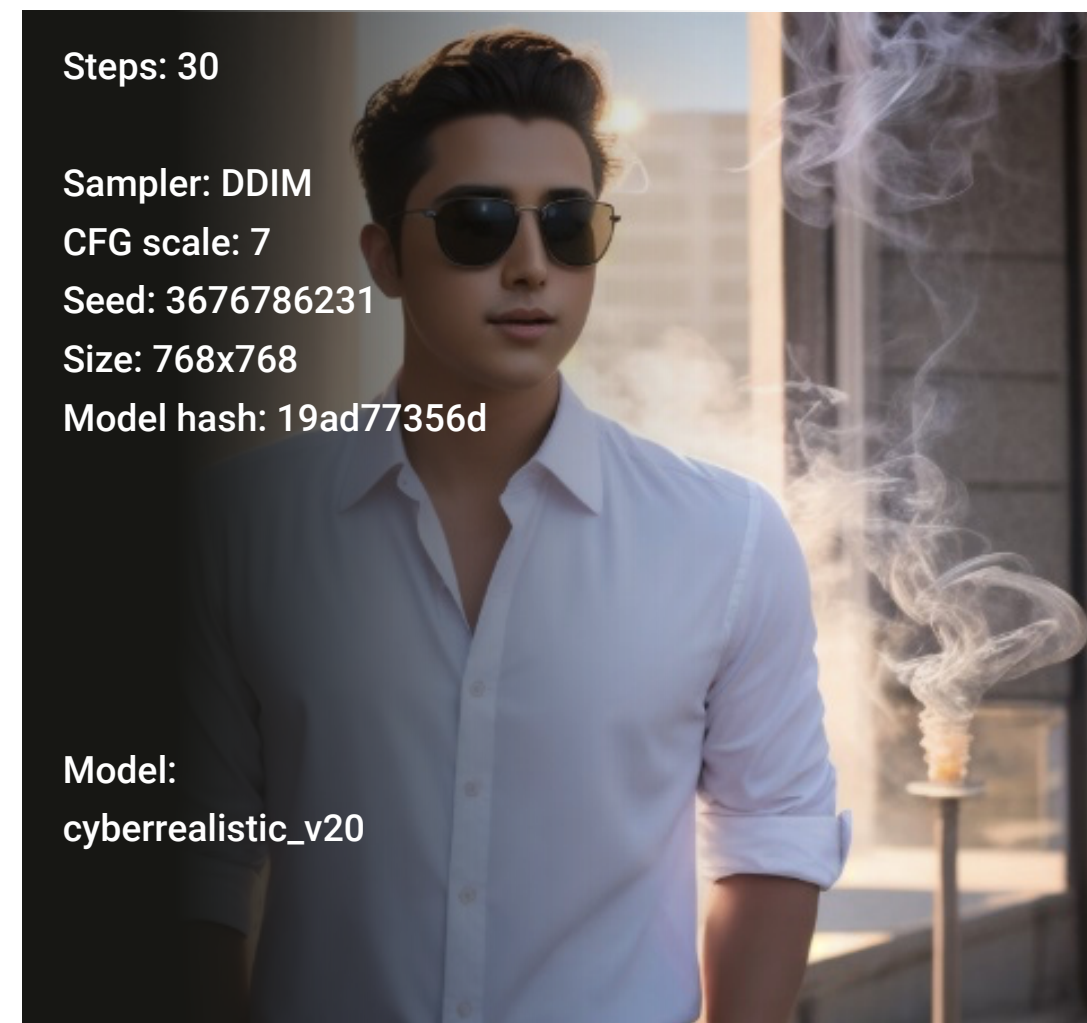


Popular Attributes

After explaining the anatomy and possible variables (or attributes), let's dive into the magical world of infinite possibilities!

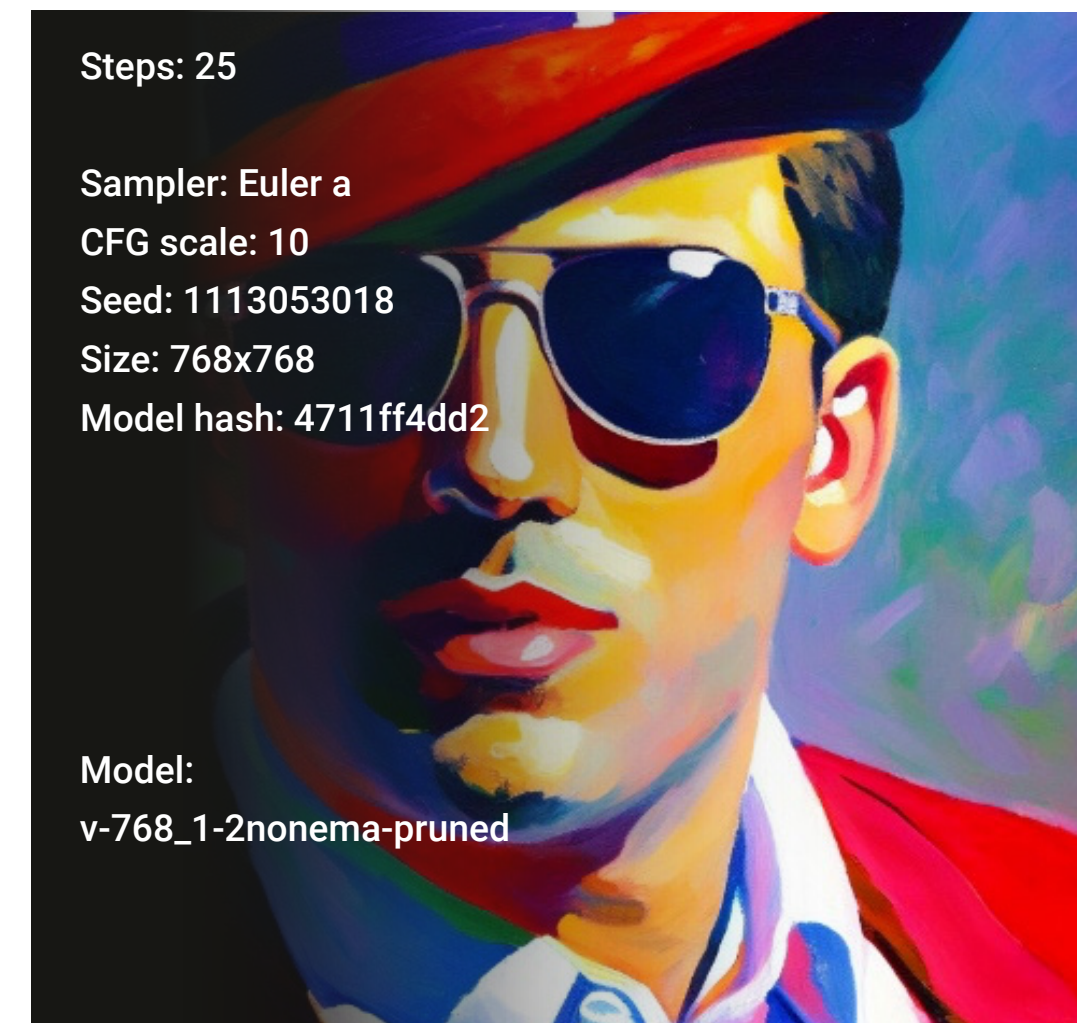
In the coming pages, we'll be exploring some of the coolest attributes.

Medium & Style



Portrait

Prompt: Photo portrait of a young man, wearing sunglasses, and smoking a cigar, cinematic lighting, 50mm, DSLR, professional photo



Painting

Prompt: Painting of a young man, wearing sunglasses, vibrant colours, artwork by Claude Monet

Inspiration Prompt

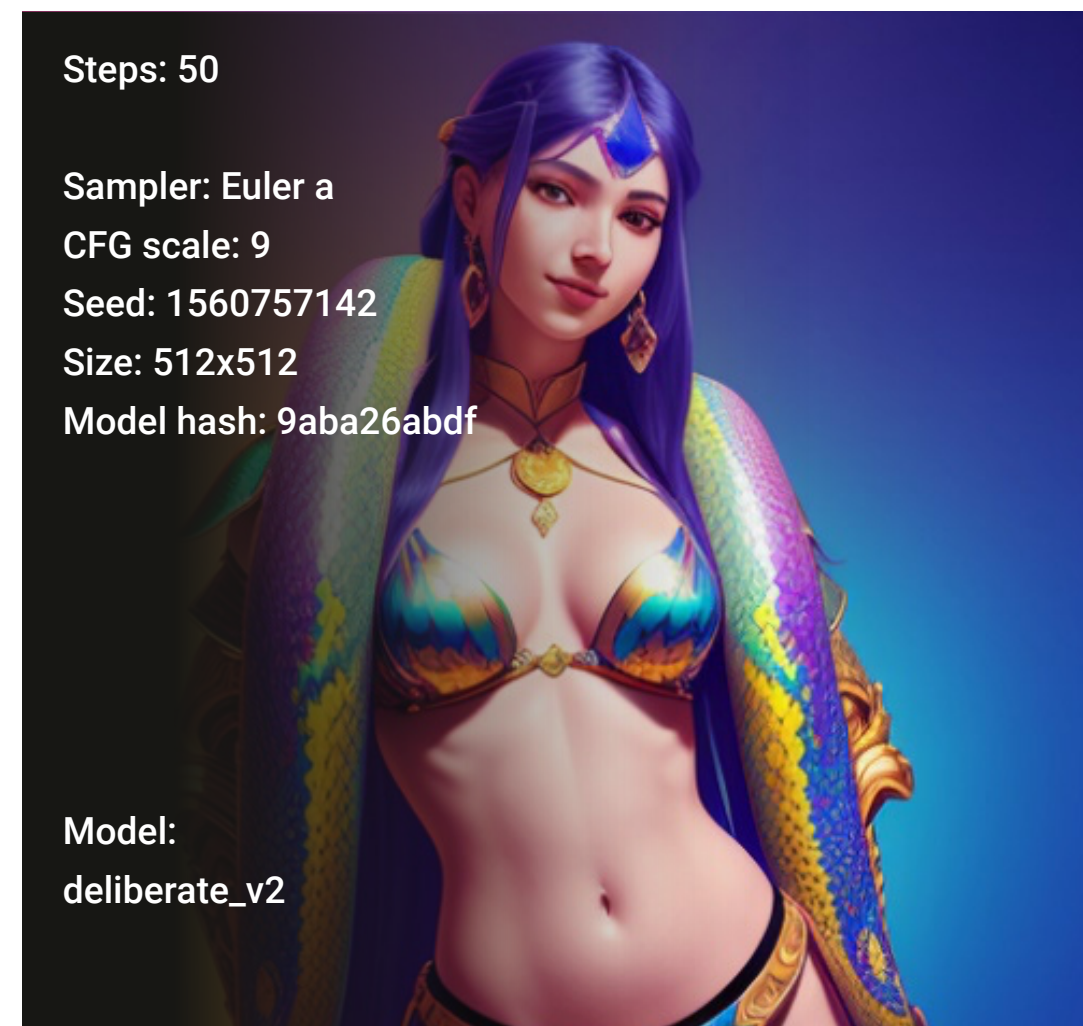


3D Art

Prompt: Pixar style little girl, 4k, 8k, unreal engine, octane render photorealistic by cosmicwonder, hdr, photography by cosmicwonder, high definition, symmetrical face, volumetric lighting, dusty haze, photo, octane render, 24mm, 4k, 24mm, DSLR, high quality, 60 fps, ultra realistic

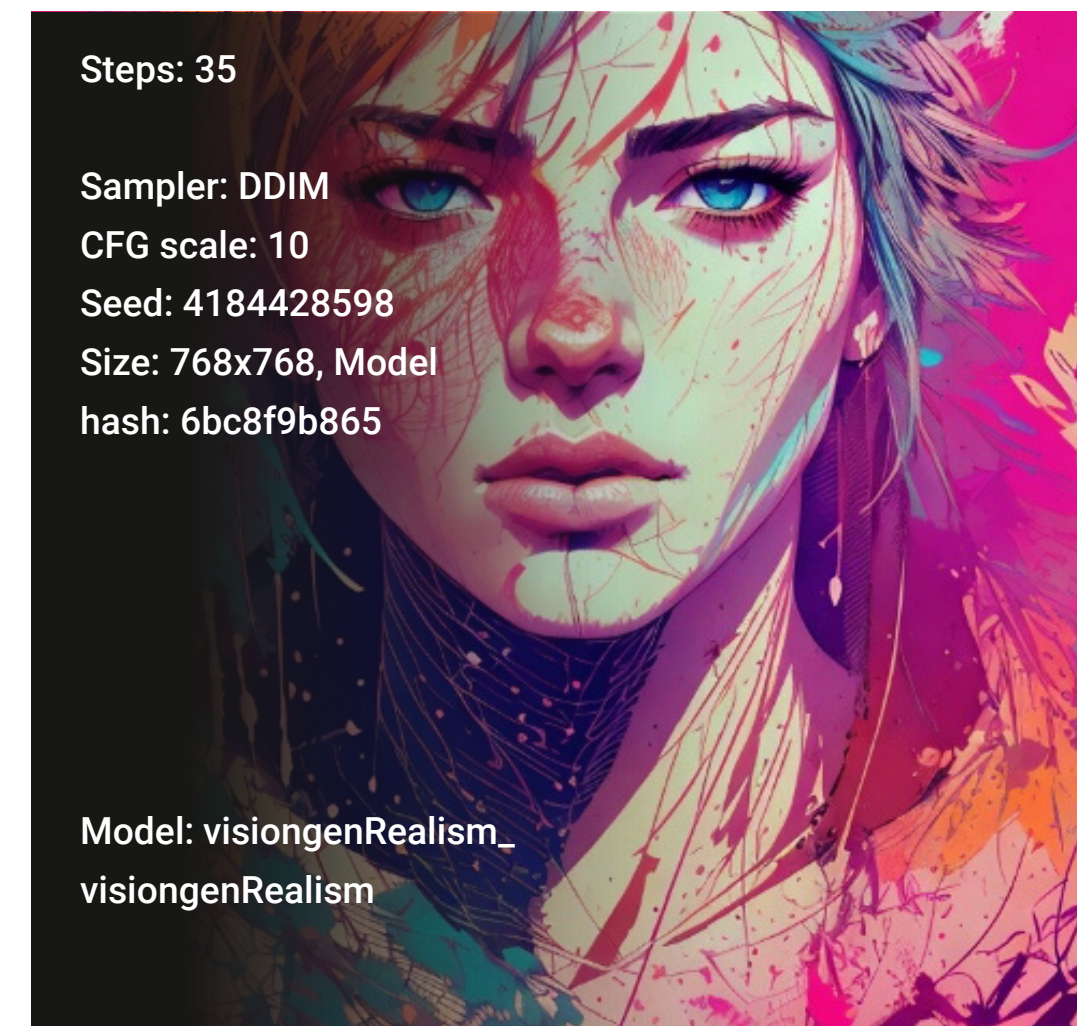
Medium & Style

Inspiration Prompt



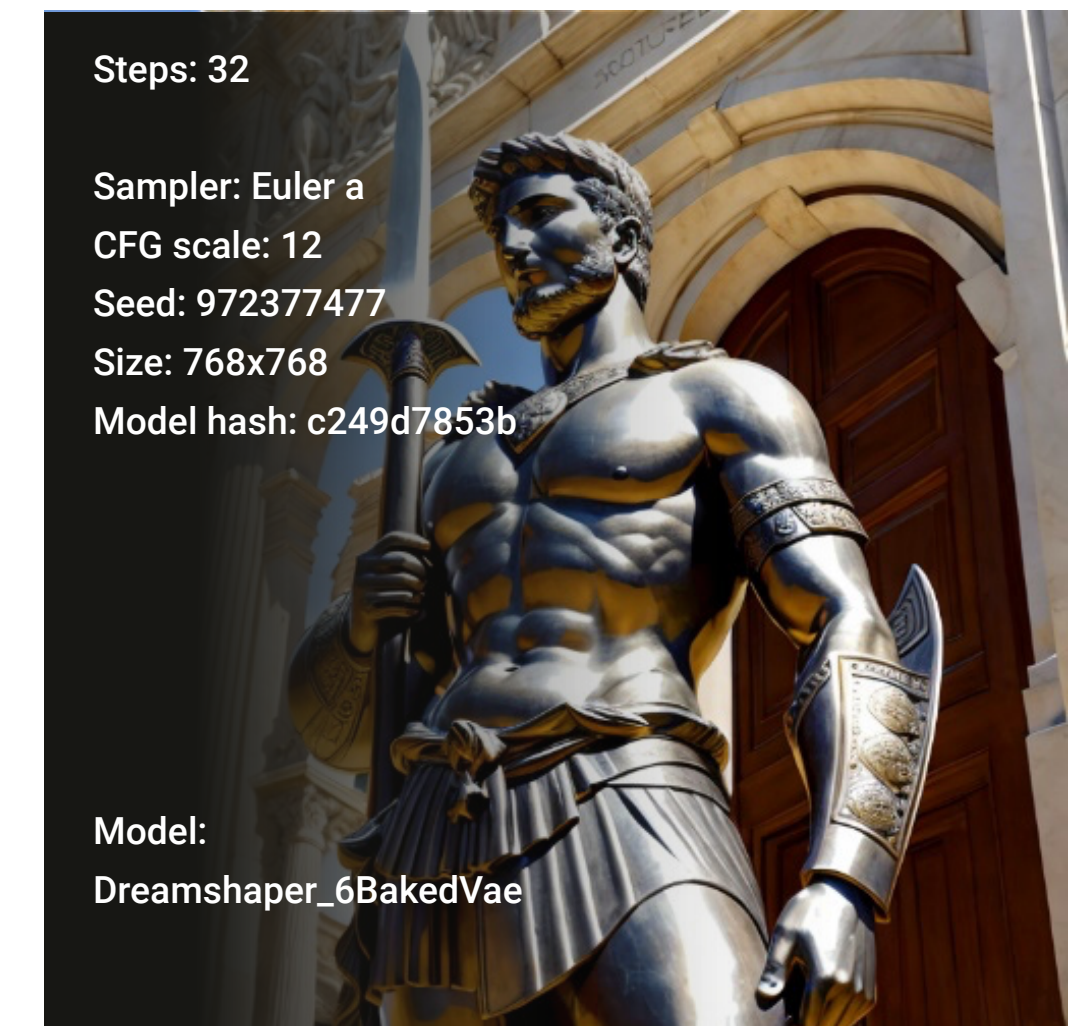
Digital Art

Prompt: Portrait of a beautiful warrior woman, ((holding a colorful viper)), smile, digital painting, illustration, au naturel, hyper detailed, digital art, trending in artstation, cinematic lighting, studio quality, smooth render, unreal engine 5 rendered, octane rendered, art by hajime sorayama



Ink

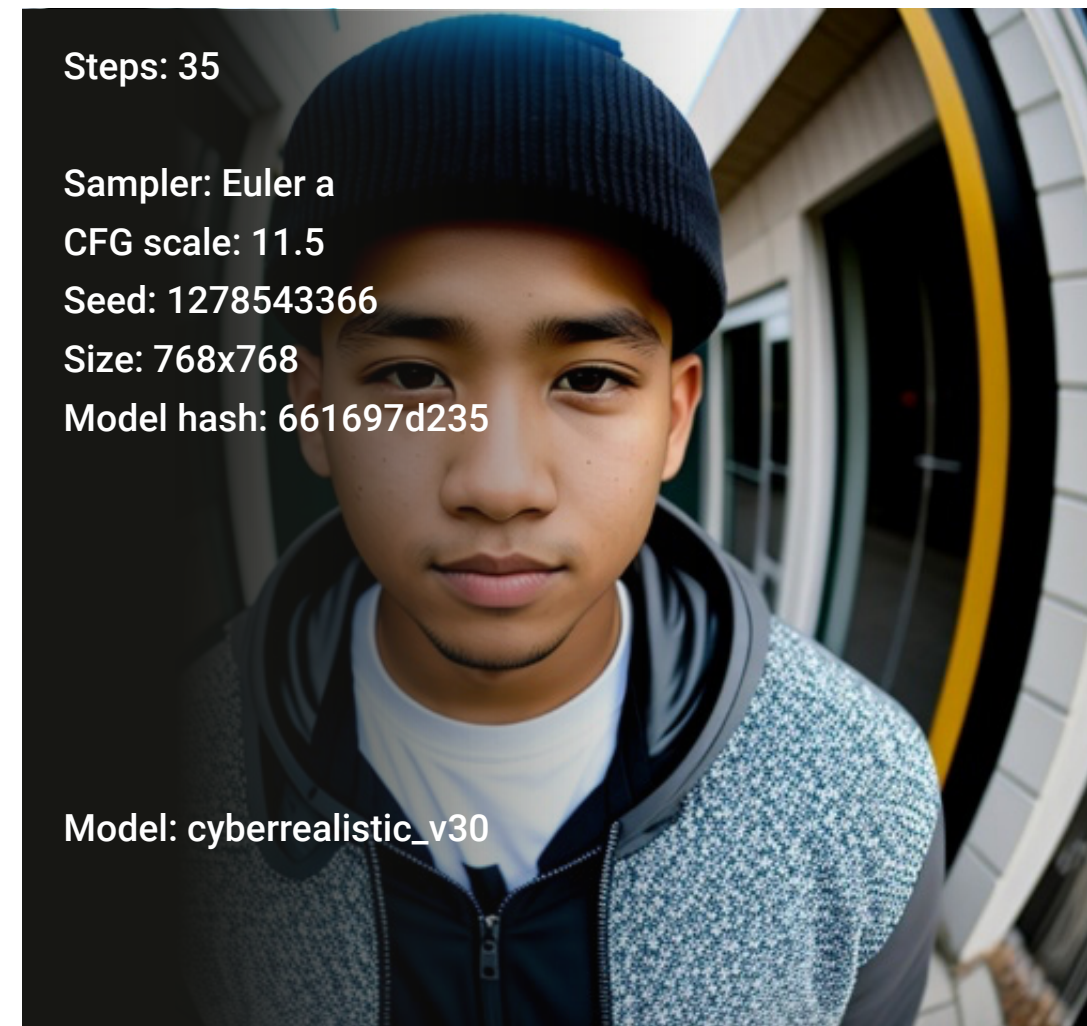
Prompt: Strong warrior prince, centered, key visual, intricate, highly detailed, breathtaking beauty, precise lineart, vibrant, comprehensive cinematic, Carne Griffiths, Conrad Roset



Sculpture

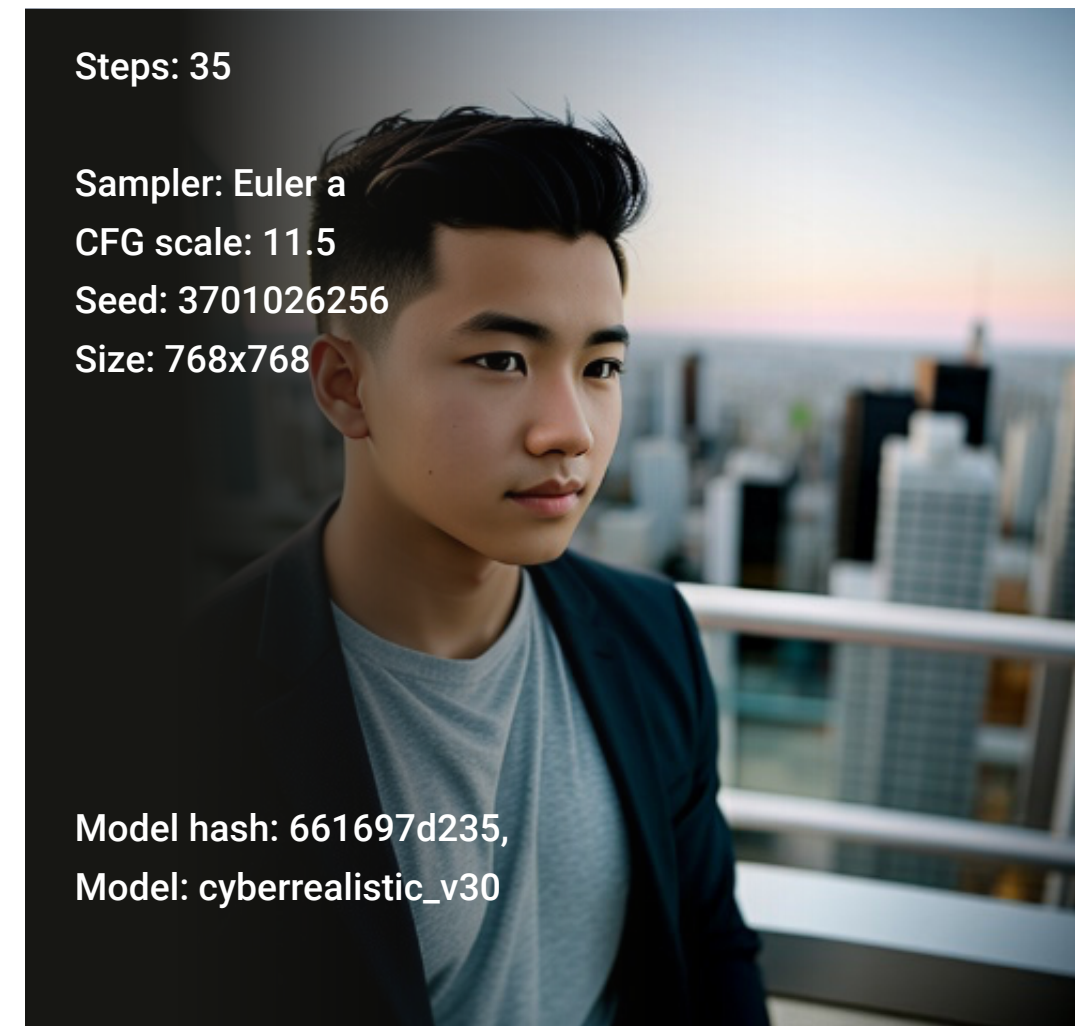
Prompt: Sculpture of a roman warrior, holding a sword, intricate detail, public street, cinematic sunlight, by Auguste Rodin quality, 60 fps, ultra realistic

Photo: Lenses



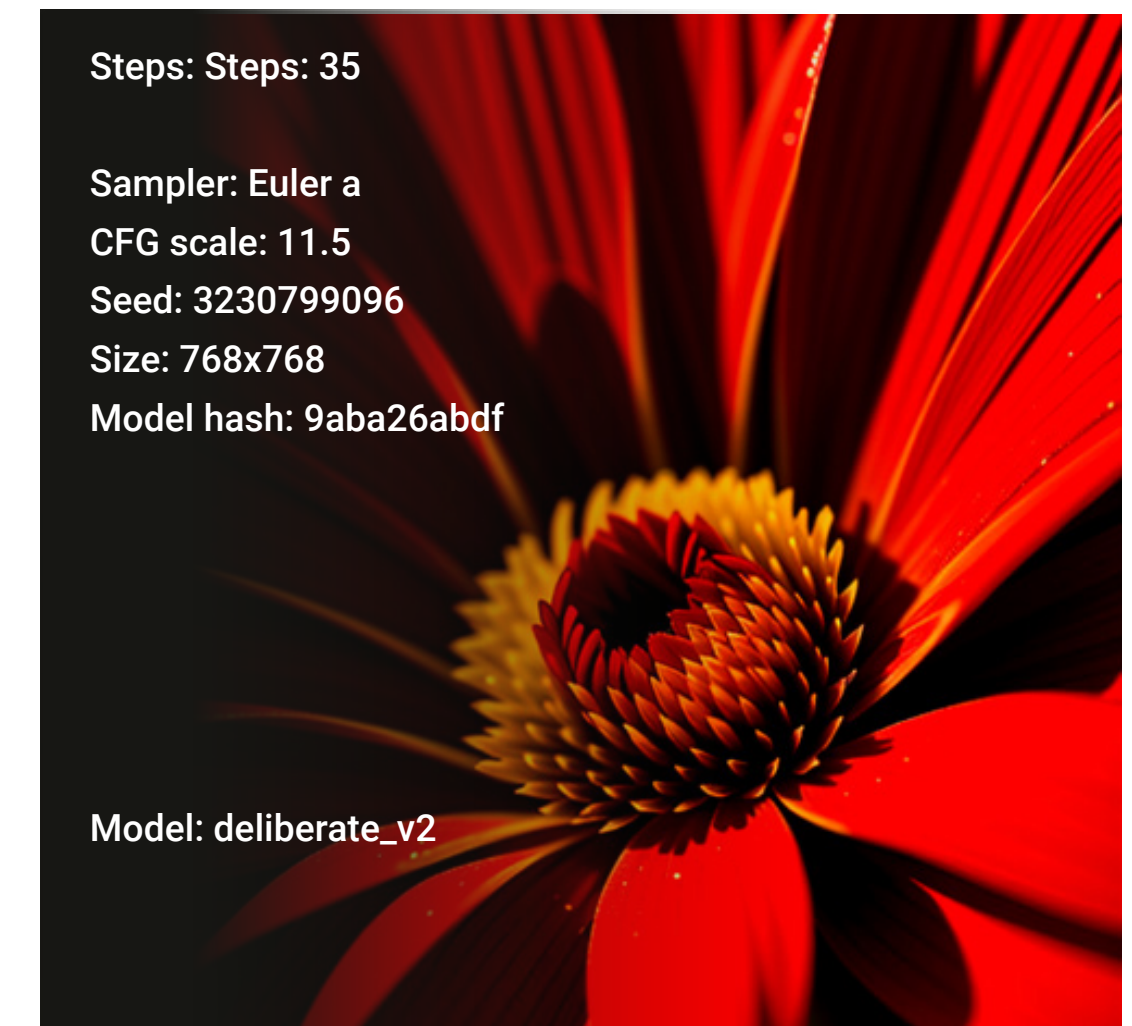
Fish-eye

Prompt: A photo, of a young man, staring into the camera lens, fisheye lens, urban clothing.



50mm

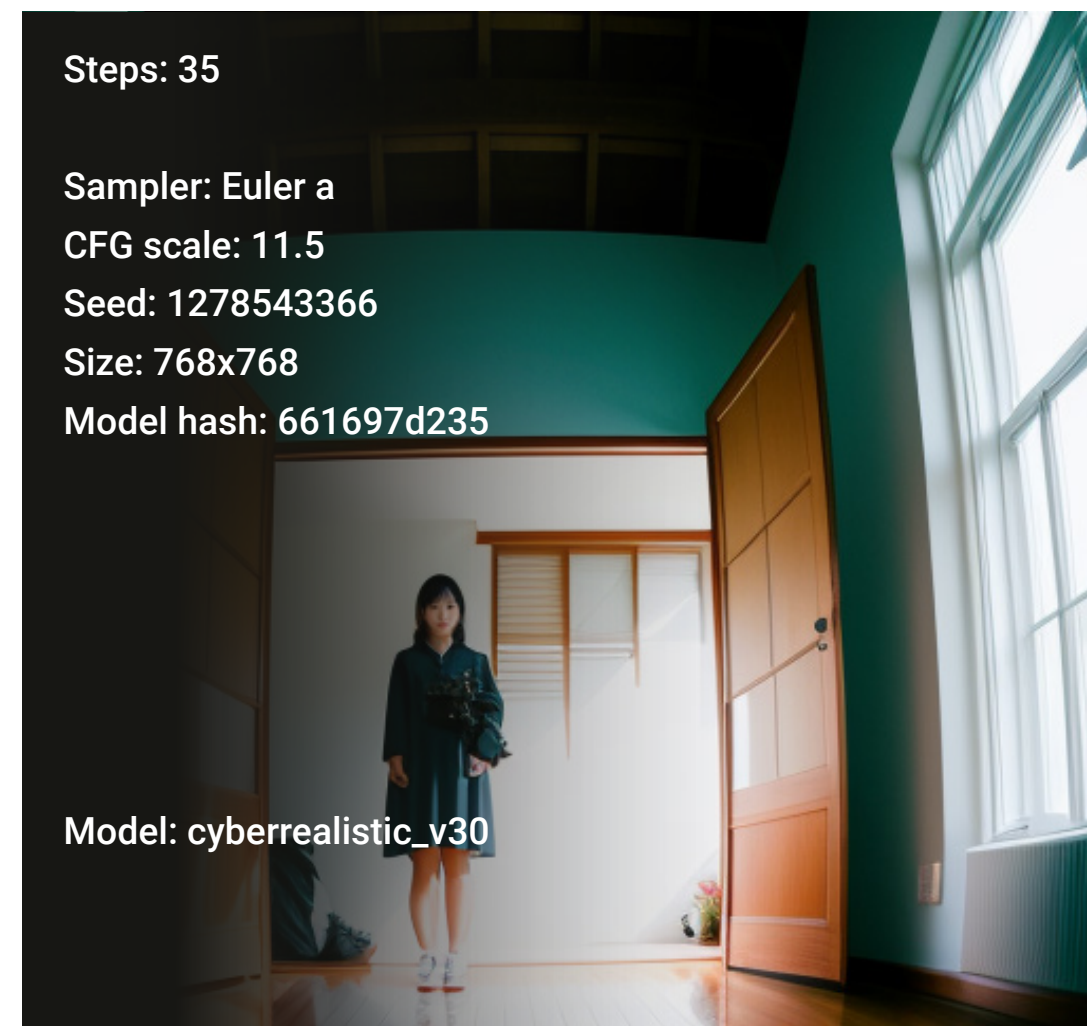
Prompt: A portrait photo, of a young man, staring into the horizon, 50mm, cityscape.



Macro

Prompt: A photo of a (red flower), macro lens, hyperrealistic, professional photograph quality, 60 fps, ultra realistic.

Photo: Focal Length



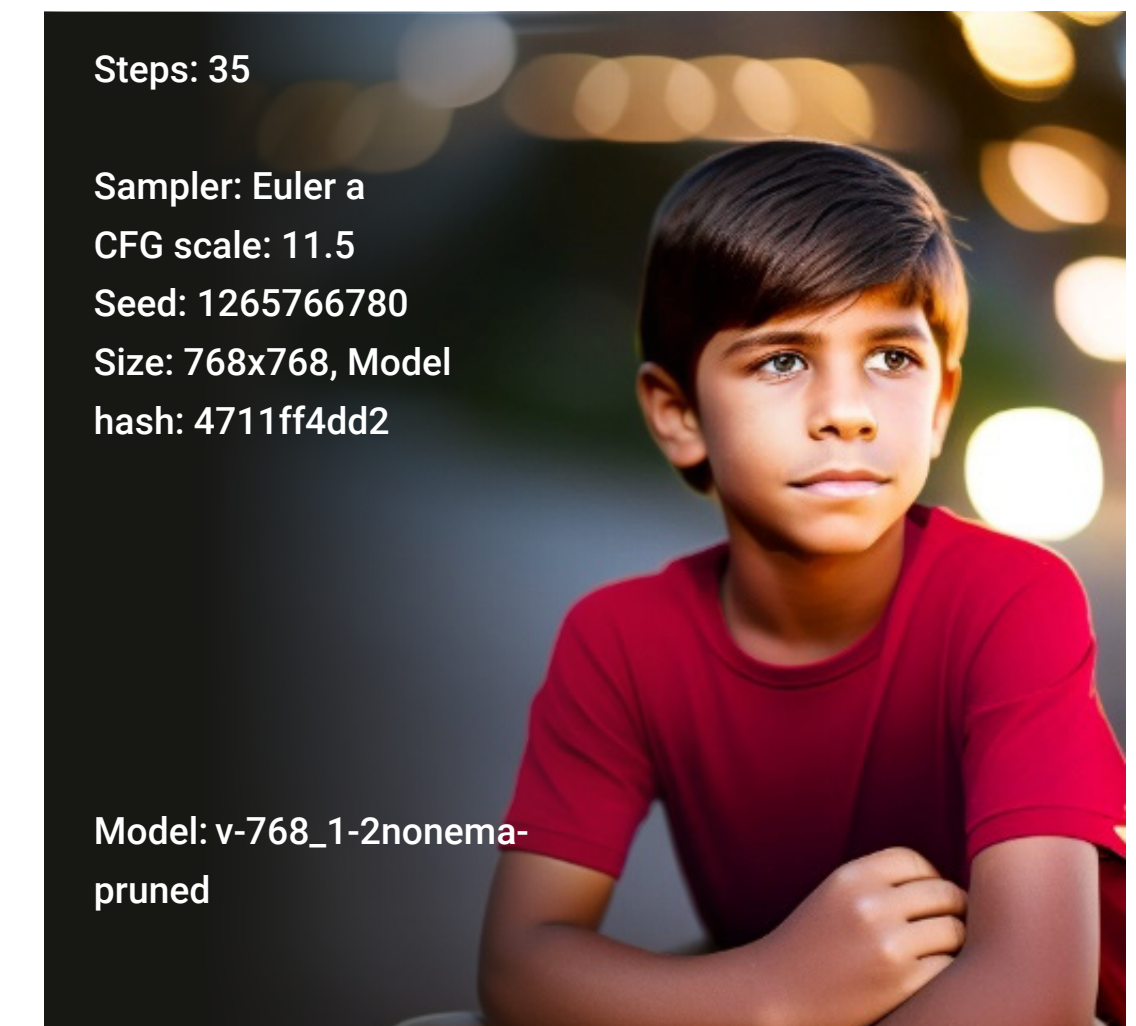
4-14mm

Prompt: A realistic photo, of an Asian woman, standing indoors, DSLR photo, (14mm lens:8 ,(1.4k, masterpiece



28mm

Prompt: A photo of a young man, DSLR, 28mm, superrealistic, detailed, hdr, masterpiece photography, cannon R5, professional photography



50mm

Prompt: A realistic photo of a young man, taken by a professional photographer, (taken with a 50mm lens:1.3), detailed, sharp focus, award winning, Steve McCurry style, realistic photo

Photo: Aperture



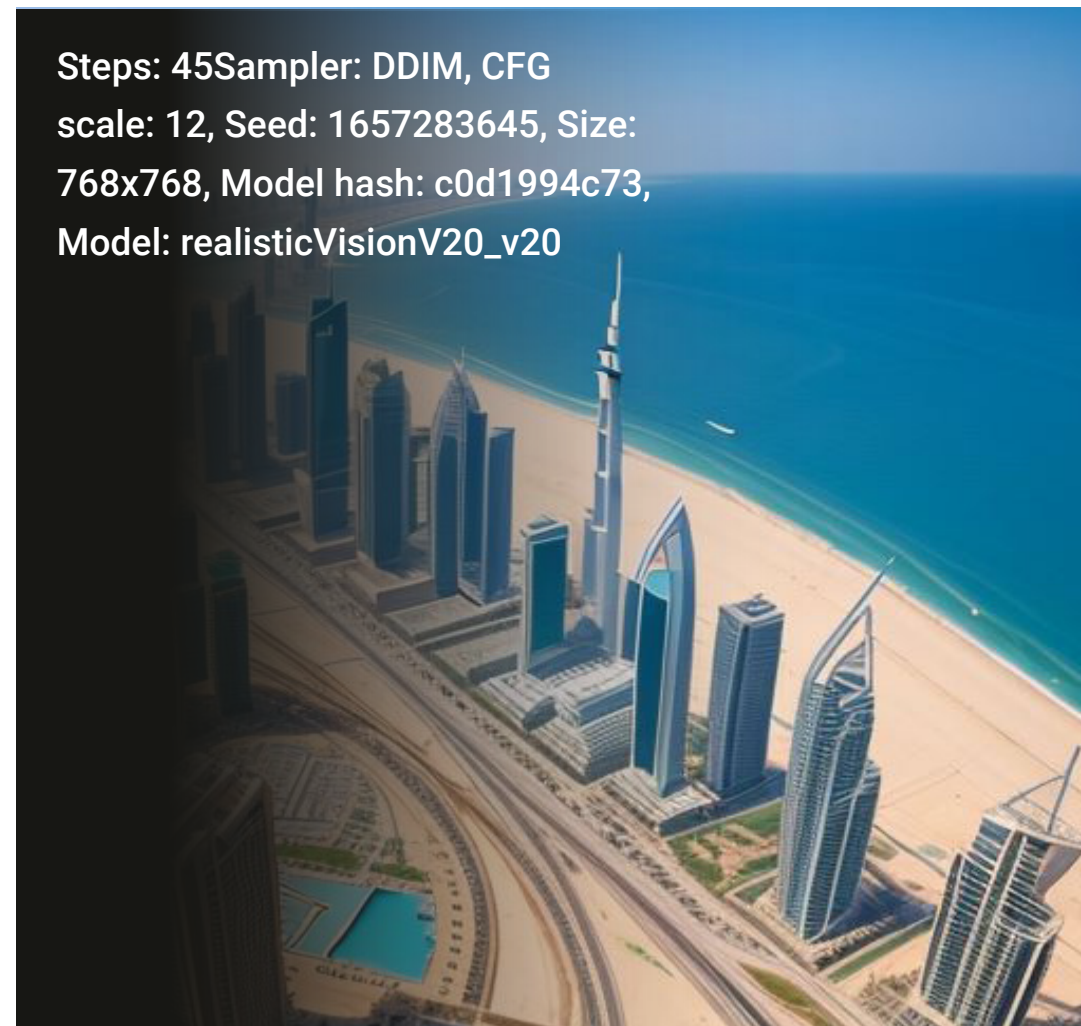
F3.4



F1.8

The aperture of a photo affects the blur you get in the background, the short term for aperture is to write “f” followed by a value, normally from 1.4 to 5 but this the possibilities are far more than that, I would recommend doing some basic research online and experimenting with different apertures, one other way to change the blur in the background would be to write “blurry background” in the prompt, or if you want a clear background, you could write “clear background” in the negative prompt section.

POV



Aerial

Prompt: Aerial shot of a city, skyscrapers, sea, sunny day, professional photo, DSLR, intricate details, focused



High-Angle

Prompt: Photo of a Buzz lightyear toy, standing on the street, high-angle shot, ultrarealistic

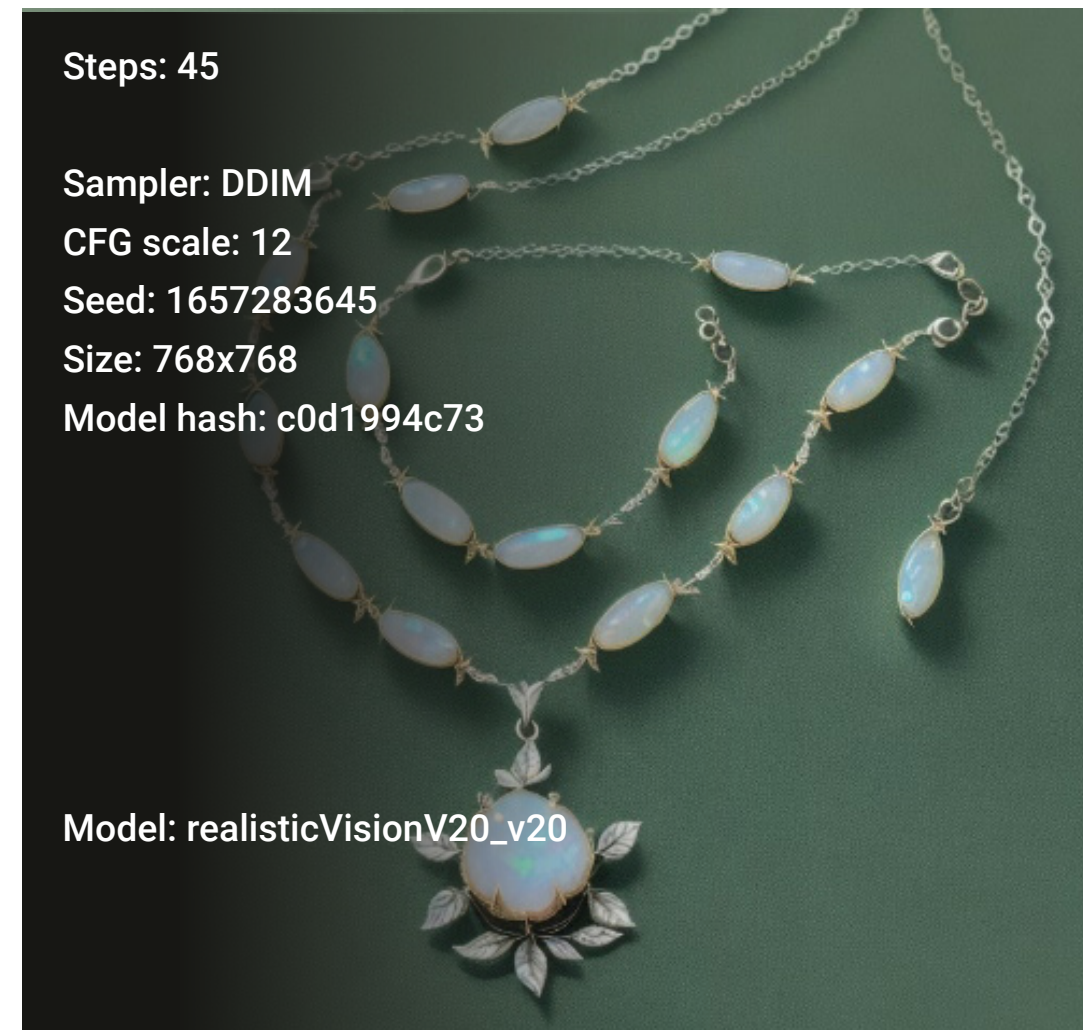


Low-Angle

Prompt: Photo of a tower, low angle shot, ultra realistic, professional photo

Subject

Source



Product

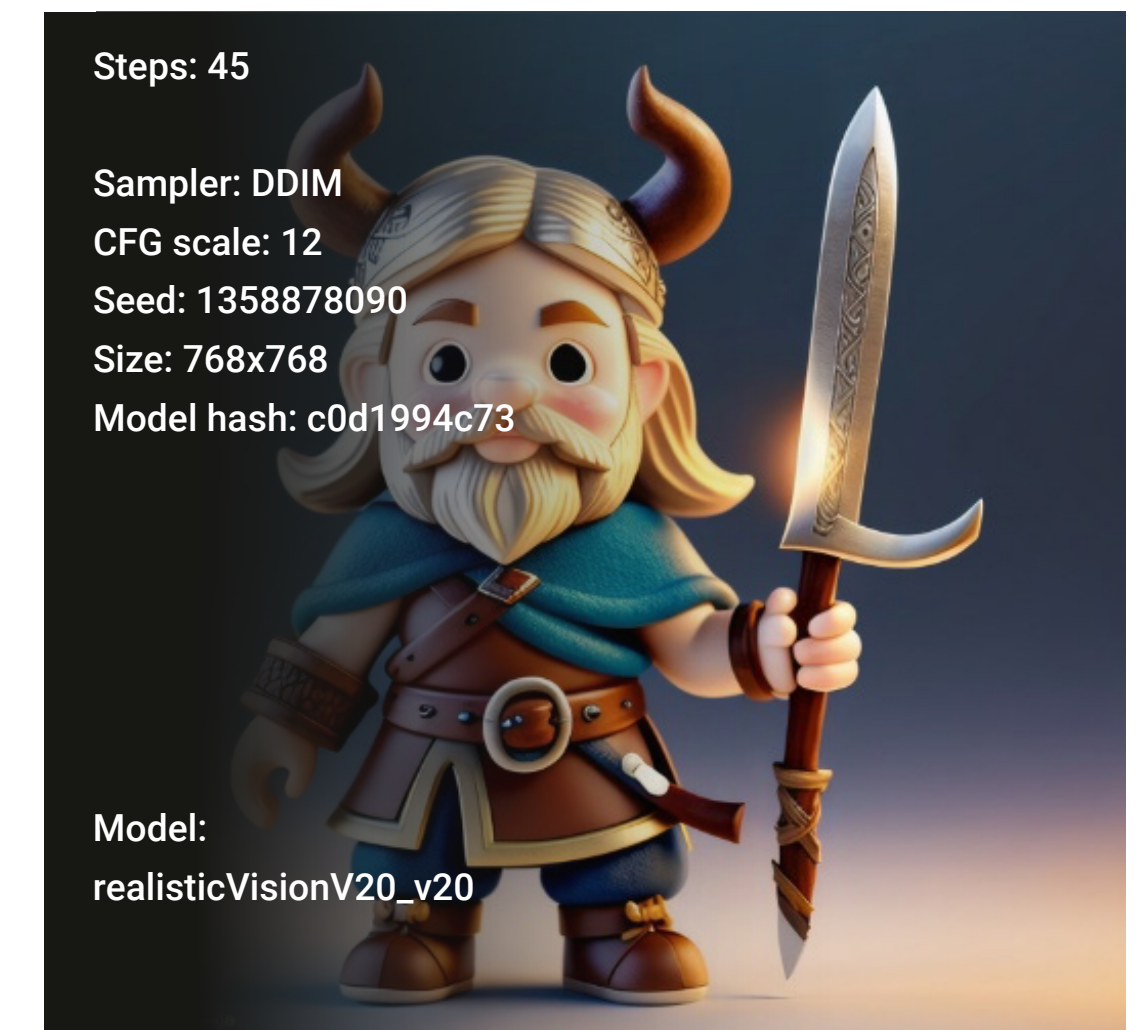
Prompt: A delicate elvish moonstone and opal necklace on a velvet background, jewelry design, symmetrical, intricate motifs, leaves, flowers, 8k octane render, high-definition photography, product photography



Animal

Prompt: A photo of a mammoth, standing in the city, surreal, photorealistic, ultra realistic, intricate detail, 8k

Prompt Source



3D Render

Prompt: Cute Disney style Viking figurine, soft light, bokeh lighting, high quality render, detailed subject, long sword, 8k

Lighting

Types of Lighting:

rim lighting, bioluminescent details, studio lighting, indirect light, halo, ektachrome, luminescence, subsurface scattering, hard shadows, glowing, soft lighting, god rays, global illumination, shimmering light, backlighting, diffused lighting, radiant light rays, volumetric lighting, caustics, specular lighting, iridescent, bloom, cinematic lighting, translucency.



Soft light

Prompt: A male pirate, photorealistic, (soft lighting:1.3), octane render, hyperrealistic, (skindentation:3 .1), (photorealistic face:8) ,(2.1k), (4k), (Masterpiece), (Best Quality), illustration, soft lighting, (specular lighting:.1 4), (high detailed skin:8 ,(2 .1k uhd, dslr, , high quality, film grain, Fujifilm XT3



Cinematic

Prompt: A photo of a modern soldier, holding a sword, walking into a street battle, cinematic lighting



Bokeh

Prompt: Pretty young girl , beautiful eyes , smiling, long hair, blue eyes, (bokeh lighting in the background:1.4), outdoors, (night time:1.3), super realistic, wide angle shot, shot on Fuji Film, detailed and realistic environment

Lighting

Lighting

rim lighting, bioluminescent details, studio lighting, indirect light, halo, ektachrome, luminescence, subsurface scattering, hard shadows, glowing, soft lighting, god rays, global illumination, shimmering light, backlighting, diffused lighting, radiant light rays, volumetric lighting, caustics, specular lighting, iridescent, bloom, cinematic lighting, translucency.



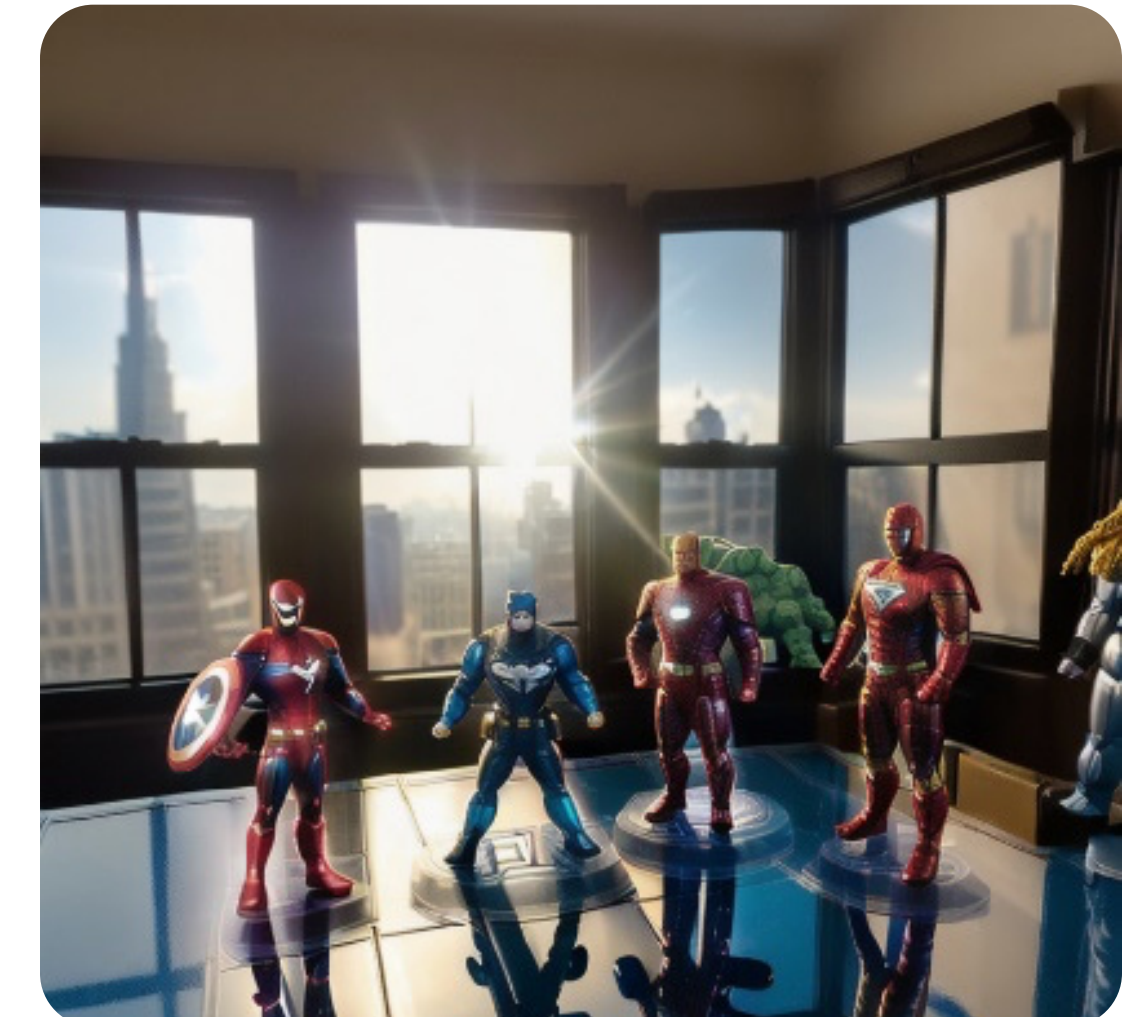
Studio Lighting

Prompt: A photo of a fashion model, professional white backdrop, (studio lighting on model's face:1.2), DSLR, f1.8, Cannon R5



Backlighting

Prompt: A photo of a man, (blue_backlight_behind the subject:1.5), DSLR, f1.8, Cannon R5



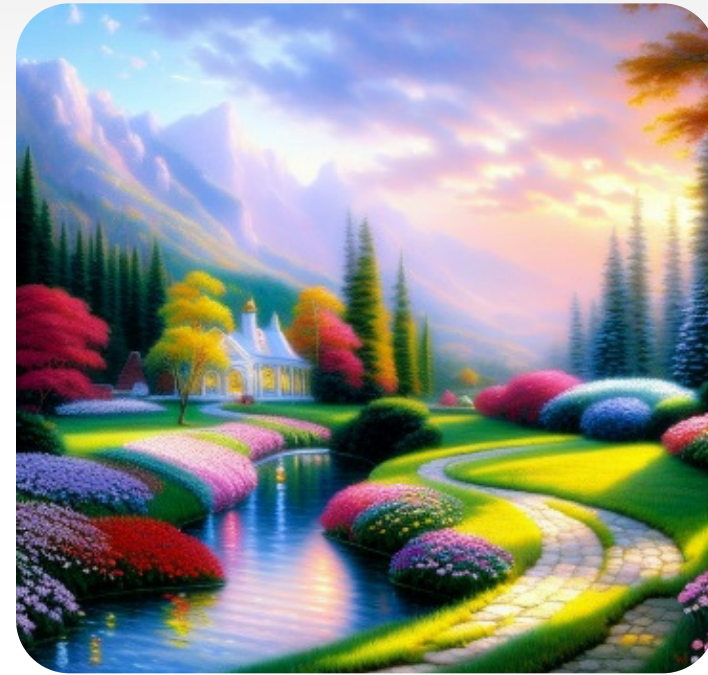
Flare

Prompt: A photo of collectible marvel and DC figurines on a table:1.4), sunlight flare, (sunlight through the window:1.1), (high quality 3D printed figurines:1.1), (entire photo is taken by a professional photographer:1.2), city view

Artists

Initially, I experimented by pairing simple prompts with the names of various artists, like "by {artist's name}". However, more often than not, the generated artworks did not resemble the particular artist's style. That was until I found a helpful article that provided a list of artists whose styles are reliably replicated by the stable diffusion model. This guide hasn't let me down yet, and here are some additional artist examples to demonstrate its efficacy. Pro tip: Always conduct some preliminary research to clarify what you're aiming to achieve before crafting your prompt. For the demonstrations in this section, I kept the prompts straightforward.

[Reference article.](#)



Thomas Kinkadee

Prompt: A painting of a heavenly utopia, by Thomas Kinkadee



Vincent Van Gogh

Prompt: A painting of a modern soldier by Vincent Van Gogh



Leonid Afremov

Prompt: A painting of Donald duck by Leonid Afremov



Claude Monet

Prompt: A portrait, of a painting of Donald trump, by Claude Monet



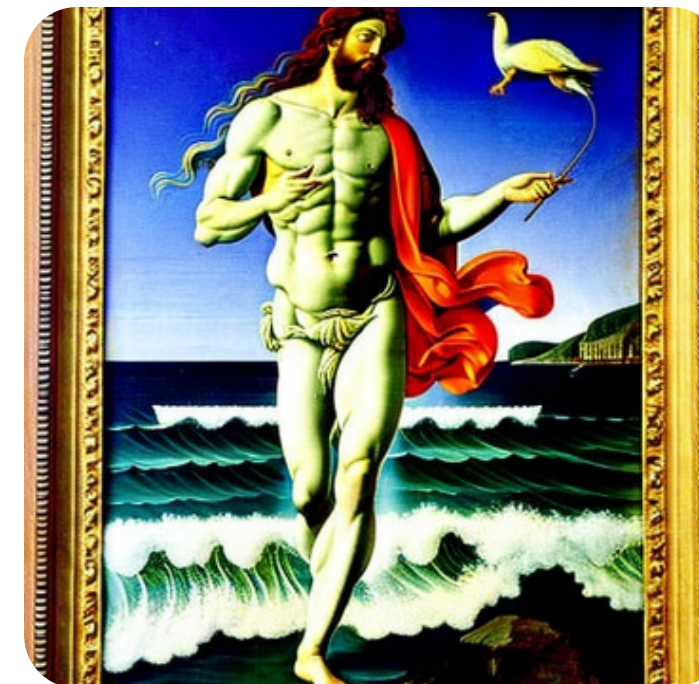
Walt Disney

Prompt: A painting of Snoop Dogg by Walt Disney



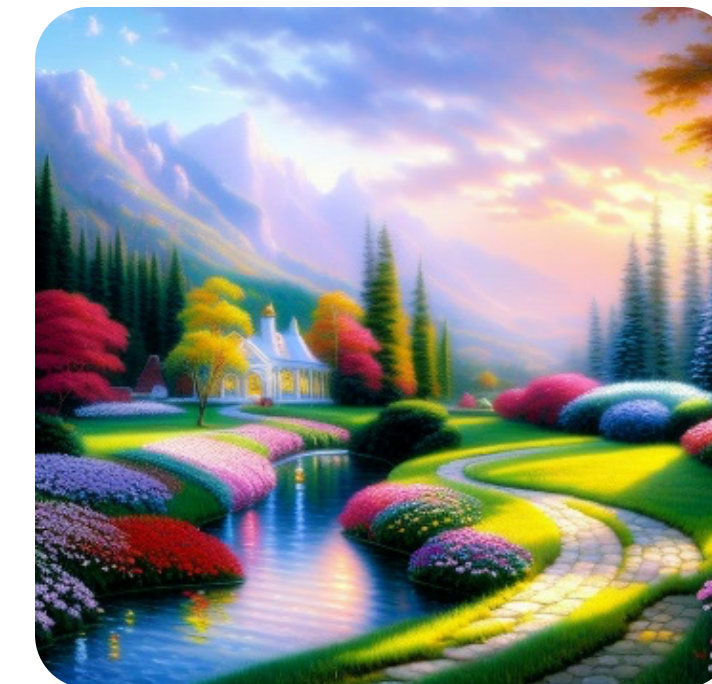
Alphonse Mucha

Prompt: (A poster of pink floyd the band:1.3), by Alphonse Mucha



Sandro Botticelli

Prompt: A painting of Poseidon by Sandro Botticelli



Kazimir Malevich

Prompt: An artwork of a (bunch of squares and triangles:1.2) by Kazimir Malevich

Websites

Source



Deviant Art

Prompt: Gal gadot, as a powerful mysterious sorceress, casting confetee magic, detailed clothing, digital painting, hyperrealistic, fantasy, Surrealist, full body, by Stanley Artgerm Lau and Alphonse Mucha, deviant art



Artstation

Prompt: Gal gadot, as a powerful mysterious sorceress, casting confetee magic, detailed clothing, digital painting, hyperrealistic, fantasy, Surrealist, full body, by Stanley Artgerm Lau and Alphonse Mucha, art station



3D Render

Prompt: Gal gadot, as a powerful mysterious sorceress, casting confetee magic, detailed clothing, digital painting, hyperrealistic, fantasy, Surrealist, full body, by Stanley Artgerm Lau and Alphonse Mucha, pinterest

Other Descriptors



RAW

Prompt: RAW candid cinema, dog, professional photo



Superrealistic

Prompt: A photo of a dog, professional photo, (superrealistic:1.3)



Trending Art

Prompt: A painting of a beautiful dog, trending art

Other Descriptors

Games



Fortnite

Prompt: A beautiful woman, holding a gun, (Fortnite game style:1.2)



League of Legends

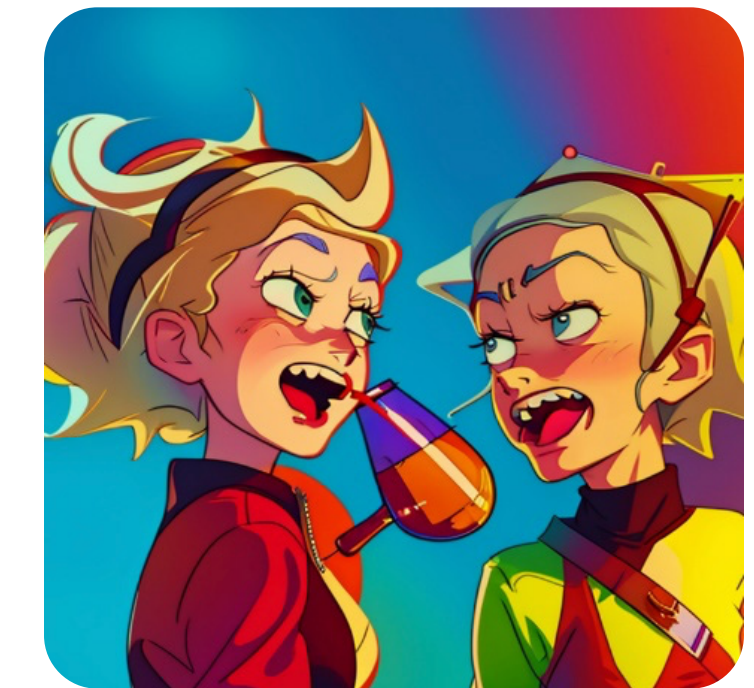
Prompt: A man walking in Cairo, (League of Legends style:1.3)

Series



Game of Thrones

Prompt: A photo of a beautiful belly dancer, young, pretty, symmetrical, public city, sunlight, (Game of Thrones scene:1.4)



Rick and Morty

Prompt: A cartoon visual of (two women laughing and drinking:1.3), radiant colors, (Rick and Morty style:1.2)

3D Descriptors

3D renders & realism

Here are some examples you might want to explore:

cinema4d, houdini render, arnold render, film photography, ultra realistic, vray, lumen reflections, bokeh, raytracing, studio quality, 8k uhd, quixel megascans, cgi, 100mm, unreal engine, cgsociety, dslr, octane render, film grain.

The visual instances provided utilized the same seed and other parameters, all encapsulated within concise prompts.



Unreal Engine

Prompt: A 3d render of Assassin's Creed, Unreal Engine



Cinema4D

Prompt: A 3d render of Assassin's Creed, Cinema 4D render



Octane Render

Prompt: A 3d render of a Assassin's Creed, Octane render

Bounus Tips



Experimenting with various variables is key; as the saying goes, “practice makes perfect.” Mastering the creation of high-quality images using Stable Diffusion can indeed be challenging. It unquestionably requires a substantial amount of practice. And remember you can mix and match styles, you can even mix and match the reference such as “in the style of Monet and Van Gogh”



Before crafting your prompt, it's advisable to conduct some research. You can conveniently utilize resources like Google or even ChatGPT to investigate prominent figures within a specific niche, such as popular directors for TV advertisements, renowned sculptors, and so forth.



Explore additional attributes or "modifiers" that can be utilized with Stable Diffusion; users are continually discovering more each day. This technology thrives on a community-rich platform, all driven by the transformative power of the collective AI revolution. Stay tuned, as I will be recommending some excellent resources towards the end!



Experiment with existing prompts that users in the community have put together, this will help you get very good results at a very early stage. There are multiple cheat sheets that can be found in the resources section at the end of this book. It is definitely recommended to tweak these prompts continuously so you can test and see what kind of results you can achieve with various keywords!



Tokens

Tokens in the context of Stable Diffusion GUI refer to the units of text used in the input prompt or text description provided to the Stable Diffusion model. When working with text-based models like Stable Diffusion, the input text needs to be divided into smaller units called tokens for processing. Tokens can be words, characters, or even subwords depending on the tokenization scheme used.

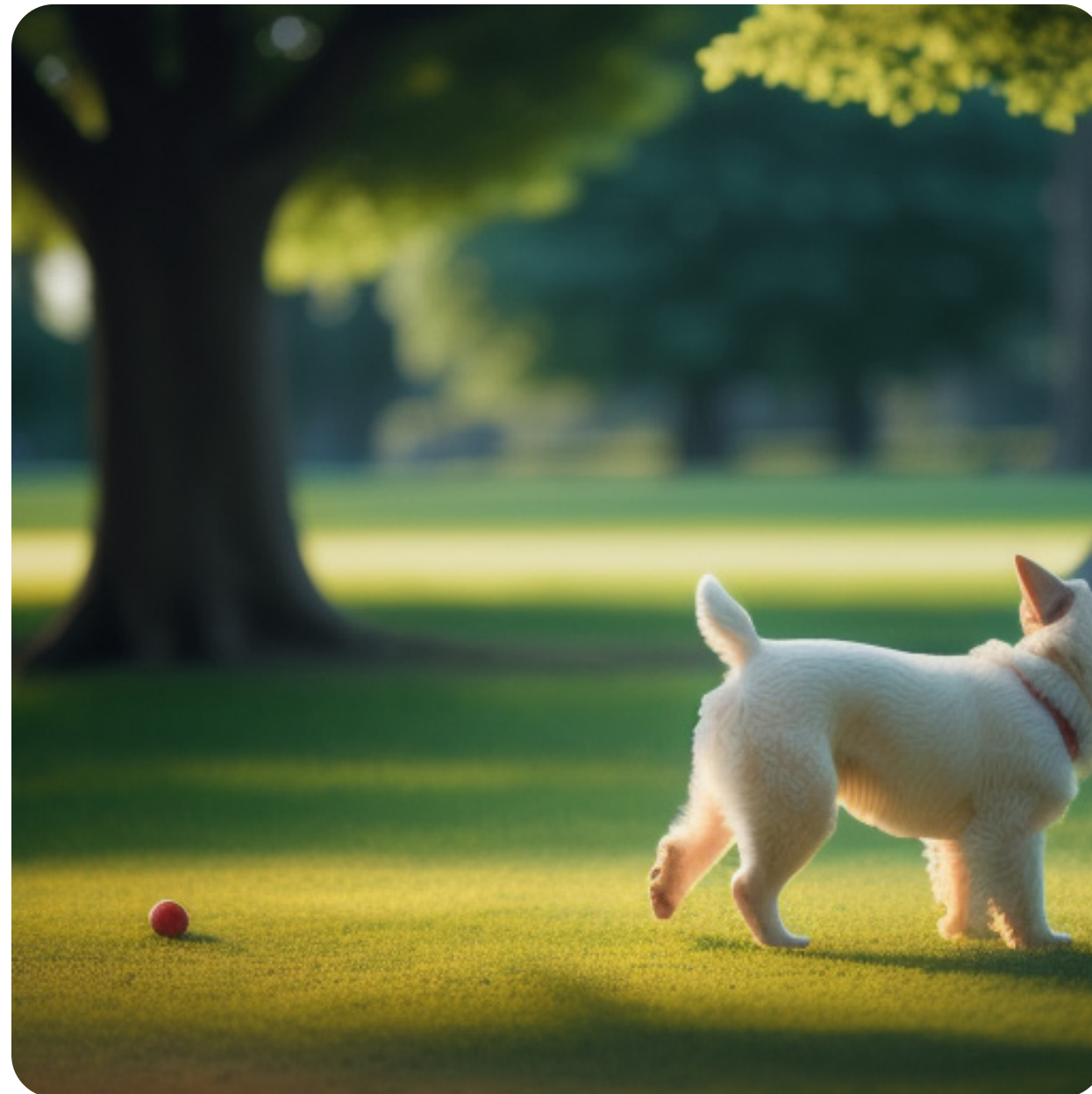
It's worth mentioning that the use of tokens is a common concept in natural language processing and deep learning models in general. Tokenization is the process of breaking down text into these smaller units to facilitate text analysis and model training. The choice of tokenization scheme can vary based on the specific requirements of the model or the task at hand.

AUTOMATIC1111 doesn't have a maximum token limit. If you use a prompt with more than 75 tokens, which is the max for the CLIP tokenizer, it simply starts a new set of 75 tokens. So, the new max becomes 150 tokens. This keeps going on until your computer runs out of memory.

Each set of 75 tokens is dealt with on its own. The results from each set are then joined together before going into Stable Diffusion's U-Net. You can see how many tokens you're using in AUTOMATIC1111. Just look at the small box in the top right corner of the box where you type in your prompt.

Word Ordering

Understand that the way you order your words in a prompt will often have a major effect on the results. This is the general rule of thumb. However, in longer prompts, the order has an even larger effect, for example; having a keyword at the very end of a long prompt, will give that keyword an extremely low weight value in the rendered image. But even on a basic level, as illustrated in the below images, just with three primary sections of the prompt “A kid”, “A dog” and “A park” the order of each prompt (although on the same seed) had entirely different final results. Try this out a few times so you can get a feel of how it works, including different variables of keywords, ordering and length in order to establish a proper idea of how you’d best order your keywords throughout your entire prompting journey.



A photo of a dog playing with a kid in the park



A photo of a kid playing with his dog in the park



A photo of a park, with a dog playing with a kid

Dimensions

512x512

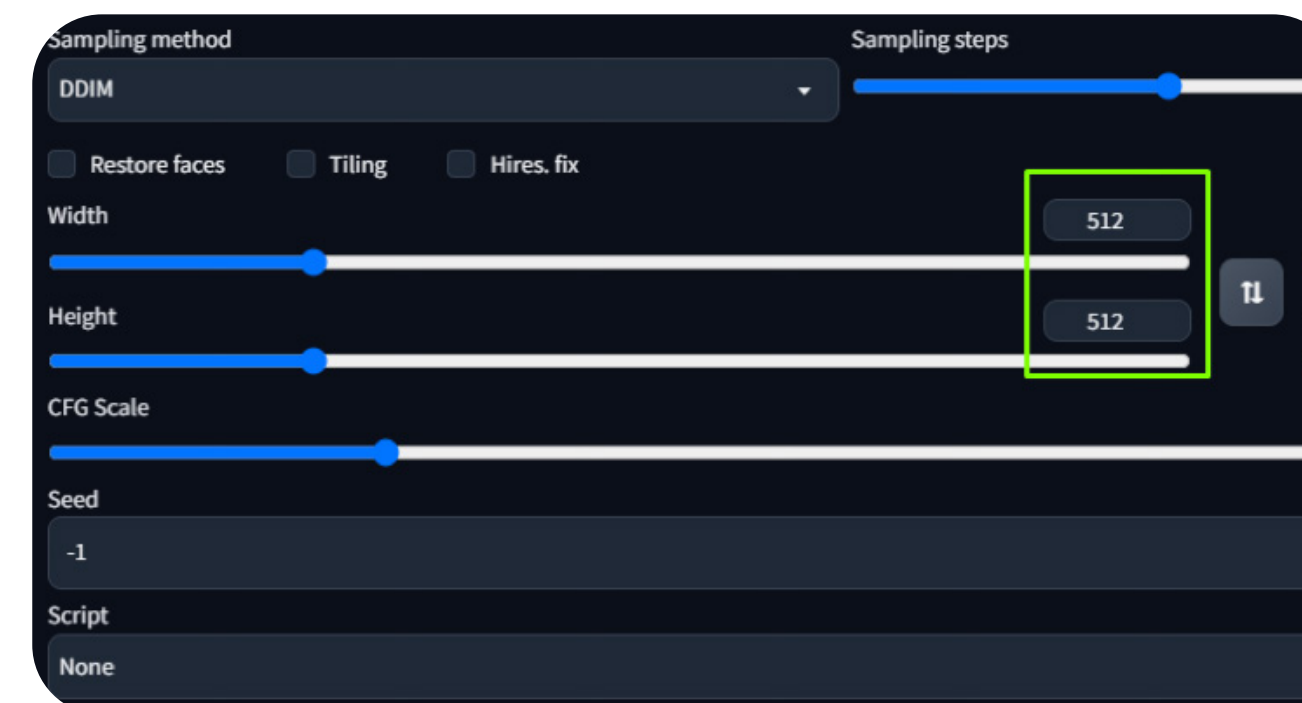
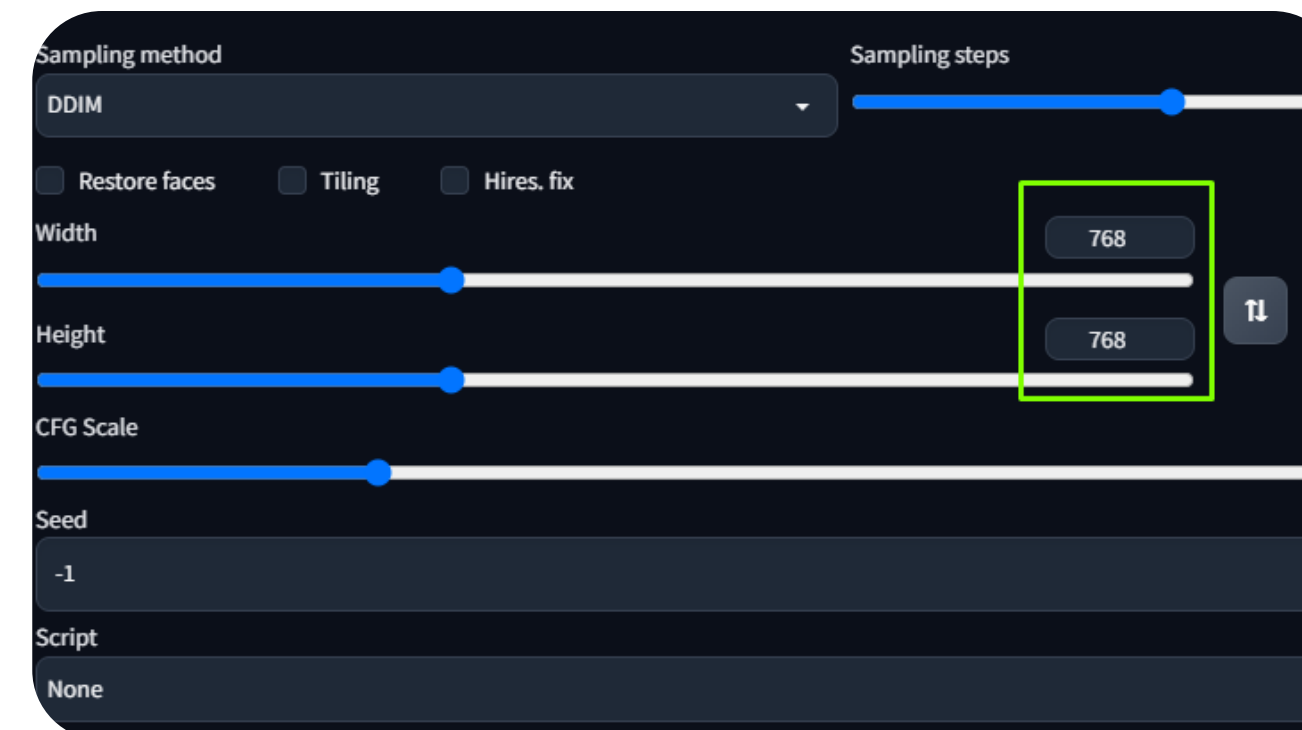
The 512x512 option is the default dimension for Stable Diffusion. It is a good choice for generating images that are of a reasonable size and quality.

768x768

The 768x768 option is a larger dimension that can be used to generate higher-quality images. However, it will take longer to generate these images and they will require more computational resources.

Other Resolutions

Although there are use cases where using other resolution parameters could work including “chunking” or “tiling”. It’s worth mentioning that this image-generation model was trained on the two options mentioned on this page, changing the resolution could result in bad output images with incoherent results, and potentially major issues with the image details.





Dimensions vs Specs

To run on 768x768, the suggested minimum Desktop spec requirements are:

Requirement	Specification
Minimum CPU	Intel Core i5 or AMD Ryzen 5
Minimum RAM	16GB
Minimum GPU	NVIDIA GeForce GTX 1060 or AMD Radeon RX 580
Minimum Storage	100GB
Suggested Internet Connection	High-speed

Optimization Tips	Close superfluous applications
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For faster or higher quality image generation, a more powerful setup may be required.

To optimize Stable Diffusion on 768x768, close superfluous applications to allocate more resources.

A high-speed internet connection aids faster image generation due to the large amount of data downloaded by the program. Remember, quality image generation takes time - patience is key.

03 | DIG DEEPER



Negative Prompting

Negative prompts serve as a tool for users to exclude specific keywords or phrases from the resulting images. These prompts play a crucial role in version 2 models, guiding the AI to avoid certain elements, thereby refining the output to align with the user's expectations. This feature provides a way for users to exert control over the image generation process. However, the availability and application of this function can vary across different versions or services. Overall, this feature enhances the quality and relevance of images, providing users with a customized creative experience.

txt2img

img2img

Extras

PNG Info

Checkpoint Merger

Train

Settings

Prompt (press Ctrl+Enter or Alt+Enter to generate)

Negative prompt (press Ctrl+Enter or Alt+Enter to generate)

Sampling method

Euler a

Sampling steps

☐ Restore faces

☐ Tiling

☐ Hires. fix

Width

512

Height

512

CFG Scale

Seed

-1

Script

None

How? You Already Know!

Getting started is very easy, all you need to do is type in keywords of what you do not want to have in your final result. Start by writing the most obvious things in the box marked in the image. You will structure your negative prompt very similarly to your positive prompt. Placing the least wanted items in keywords at the beginning of the prompt.

With vs Without

To illustrate the impact, I conducted a brief experiment, using the same prompt twice: once with and once without a negative prompt. Can you guess which image included a negative prompt to refine the result?

We must consider that crafting clear and thoughtful prompts can produce high-quality images. However, negative prompts remain crucial for refining your output. Moreover, they have become even more effective V2.1.



With vs Without

A large part of the Image-generation and prompting experience revolves around problem-solving, and one of the primary methods of solving image mistakes is to add the right keywords that will have an immediate impact on the next image you generate, in different contexts, different solutions or many different solutions may be required, but in others, one single keyword can make all the difference.



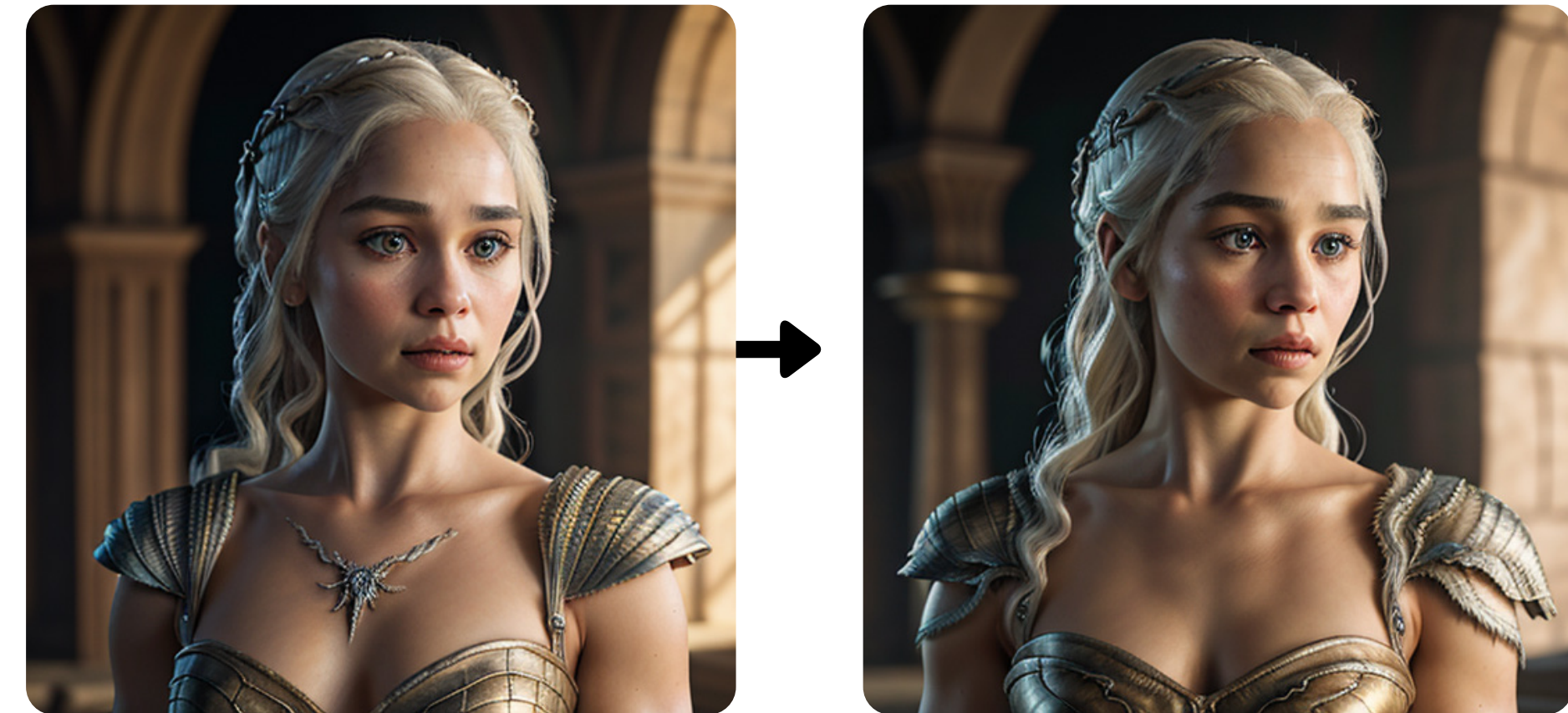
Added the negative prompt keyword: (((tiling)))
and made no other changes

With vs Without

Pro Tip:

If you're aiming to make subtle modifications to a generated image, start by copying the "seed" which you'll find at the bottom-right of the GUI details attached to the original image. Insert this seed into your seed option, then supplement it with a negative prompt before executing the re-render. As an example, I used this method to strip Daenerys Targaryen of her jewelry. While inpainting could be a more refined solution to ward off minor alterations during the second render, this is a swift and effortless strategy that all AI image enthusiasts and artists should have in their toolbox. This can also be achieved with inpainting, and perhaps that would be the better option to avoid even minor changes that occurred when rendering my image a second time, as illustrated in these images.

In this example, I added the negative prompt keywords: ((neck jewelry)) and made no other changes to the prompt or GUI settings.



With vs Without

You can also change entire aspects of a scene with one word.

The first image was generated without me adding any words to imply I want to have a green front yard, now considering I didn't want that yard, or the overall greenery, I then went through a couple of tries with negative prompts, first with one keyword then with three separate ones, to achieve the result I was looking for.



((green))



((trees)), ((greenery)), ((green))



Utilize Other User Suggestions

In most cases, you will easily find negative prompts online which you can copy and paste into your GUI, then tweak and adjust according to your specific needs.

Here are some popular sources you should check out:

[Github](#)

[Hugging Face](#)

[Nerdschalk](#)

[Sebastian Kamph's presets](#)

Community Examples

Source1 Source2 Source3 Source4

You can copy paste these examples and try them out for each corresponding use-case:

(deformed, distorted, disfigured:1.3), poorly drawn, bad anatomy, wrong anatomy, extra limb, missing limb, floating limbs, (mutated hands and fingers:1.4), disconnected limbs, mutation, mutated, ugly, disgusting, blurry, amputation

(deformed, distorted, disfigured:1.3), bw, doll, drawing, cartoon, painting, illustration, graphic, cgi, poorly drawn, bad anatomy, wrong anatomy, extra limb, missing limb, floating limbs, (mutated hands and fingers:1.4), disconnected limbs, mutation, mutated, ugly, disgusting, blurry, amputation

cropped head, black and white, slanted eyes, deformed, bad anatomy, disfigured, poorly drawn face, mutation, mutated, extra limb, ugly, disgusting, poorly drawn hands, missing limb, floating limbs, disconnected limbs, malformed hands, blurry, (((mutated hands and fingers))), watermark, watermarked, oversaturated, censored, distorted hands, amputation, missing hands, obese, doubled face, double hands

(((((ugly))))), (((duplicate))), ((morbid)), ((mutilated)), out of frame, extra fingers, mutated hands, ((poorly drawn hands)), ((poorly drawn face)), (((mutation))), (((deformed))), ((ugly)), blurry, ((bad anatomy)), (((bad proportions))), ((extra limbs)), cloned face, (((disfigured))), out of frame, ugly, extra limbs, (bad anatomy), gross proportions, (malformed limbs), ((missing arms)), ((missing legs)), (((extra arms))), (((extra legs))), mutated hands, (fused fingers), (too many fingers), (((long neck))), (((ugly))), (((duplicate))), ((morbid)), ((mutilated)), out of frame, extra fingers, mutated hands, ((poorly drawn hands)), ((poorly drawn face)), (((mutation))), (((deformed))), ((ugly)), blurry, ((bad anatomy)), (((bad proportions))), ((extra limbs)), cloned face, (((disfigured))), out of frame, ugly, extra limbs, (bad anatomy), gross proportions, (malformed limbs), ((missing arms)), ((missing legs)), (((extra arms))), (((extra legs))), mutated hands, (fused fingers), (too many fingers), (((long neck))), (((ugly))), (((duplicate))), ((morbid)), ((mutilated)), out of frame, extra fingers, mutated hands, ((poorly drawn hands)), ((poorly drawn face)), (((mutation))), (((deformed))), ((ugly)), blurry, ((bad anatomy)), (((bad proportions))), ((extra limbs)), cloned face, (((disfigured))), out of frame, ugly, extra limbs, (bad anatomy), gross proportions, (malformed limbs), ((missing arms)), ((missing legs)), (((extra arms))), (((extra legs))), mutated hands, (fused fingers), (too many fingers), (((long neck)))

canvas frame, cartoon, 3d, ((disfigured)), ((bad art)), ((deformed)), ((extra limbs)), ((close up)), ((b&w)), wierd colors, blurry, (((duplicate))), ((morbid)), ((mutilated)), [out of frame], extra fingers, mutated hands, ((poorly drawn hands)), ((poorly drawn face)), (((mutation))), (((deformed))), ((ugly)), blurry, ((bad anatomy)), (((bad proportions))), ((extra limbs)), cloned face, (((disfigured))), out of frame, ugly, extra limbs, (bad anatomy), gross proportions, (malformed limbs), ((missing arms)), ((missing legs)), (((extra arms))), (((extra legs))), mutated hands, (fused fingers), (too many fingers), (((long neck))), Photoshop, video game, ugly, tiling, poorly drawn hands, poorly drawn feet, poorly drawn face, out of frame, mutation, mutated, extra limbs, extra legs, extra arms, disfigured, deformed, cross-eye, body out of frame, blurry, bad art, bad anatomy, 3d render

lowres, bad anatomy, bad hands, text, error, missing fingers, extra digit, fewer digits, cropped, worst quality, low quality, normal quality, jpeg artifacts, signature, watermark, username, blurry



Prompt Punctuation

To craft the perfect prompt, it is essential to comprehend the distinct influence of punctuation marks and symbols. While the impact can slightly vary depending on the model and its version, the fundamental principles generally hold true. By deploying suitable punctuation and properly structuring your prompt, you can significantly enhance the output. The key punctuation marks to bear in mind are parentheses, square brackets, commas, and periods.



Commas

Think of commas as a helping hand guiding Stable Diffusion, the AI model, in better understanding the essence of your visual instructions. By using commas to break up the main elements of your image request, you'll find the output to be more precise. Let's consider a case where you want a picture of *"A lion, wearing a red cape, resting on a rock."* This phrasing tends to yield a more satisfying result than if you were to simply state "A lion wearing a red cape resting on a rock," without any commas.



Brackets (Parentheses)

NOTE: Brackets are the second most commonly used punctuation mark after commas! Using brackets is like whispering specific instructions to the model about which elements should go together or in what order things should be displayed. For instance, let's consider "A dog (wearing a bandana) (lying on a mat)." This is likely to produce a more accurate image of a dog sporting a bandana and comfortably resting on a mat than the phrase "A dog wearing a bandana lying on a mat."

Later in this book, we'll delve into a more detailed exploration of how to use brackets for customizing tokens and *weights*.

Vertical Bars

Important note: This works differently if you're not using the Automatic1111 GUI

You use vertical bars to specify different aspects of an image, such as the color, size, or pose of the subject. For example, the following prompt would generate an image of a large, blue cat sitting in a chair:

A large blue cat sitting in a chair | oil painting



Periods

Now, periods come into play when you want to finalize your image prompt. They serve as a way to sign off your request, clarifying the overarching theme for the model. For instance, if you propose, "A parrot wearing a pirate hat." the image created is likely to be more on-point compared to when you give the prompt "A parrot wearing a pirate hat" without any period in the end.

From my experience, using periods between tokens or words can have the same effect as commas.



Remember

It's like you're gently guiding your AI friend here to create the visual feast you have in mind. By providing clear and punctuated instructions, you're helping it to serve you better. It's just like communicating with a friend - the clearer you are, the better they'll understand you.

The secret here is to make your instructions clear and specific, just like how you'd communicate with a dear friend. The more detailed you are, the better the model can understand and generate the masterpiece you have in mind. Also, you must always rely on your own experimentation, the more you try things out and experience the differences between different parameter scales, wording and prompt punctuation, the more you will learn and master the craft.



Visual inspired by Alex Grey



Prompt Weights

Prompt weights serve as a guiding mechanism in Stable Diffusion, allowing you to dictate the prominence of various keywords in your prompts. By employing prompt weights, you effectively signal the model that certain keywords should carry more significance, which can prove beneficial in producing more detailed images or images with desired attributes. We will delve into two methods of applying prompt weights in Stable Diffusion in the subsequent segment.



The following table offers a concise guide to using syntax to control emphasis in Stable Diffusion.

Parentheses () are used to increase the emphasis or weight of a keyword, effectively telling the model that the enclosed keyword is of higher importance.

On the contrary, square brackets [] are used to decrease the emphasis or weight of a keyword.

Curly brackets {} represent NAI’s alternative to using parentheses for the same purpose of emphasis.

Angle brackets <> have a specific purpose for embedding usage.

Furthermore, the colon serves as a shortcut to specify multiple parentheses in a more streamlined manner. For instance, (Blue hair) would be given more importance than [Blue hair] in the final output of the model.

Moreover, (blue hair:1.4) would further increase the emphasis on ‘blue hair’ by about %40, while (blue hair:0.6) would decrease it by the same amount.

Syntax	Description
()	Increases emphasis
[]	Decreases emphasis
{}	NAI's alternative for ()
< >	Used for embeddings
:	Used to specify the number of ()s, preventing the need for multiple ()s
(Blue hair)	Gives ‹Blue hair› more weight in the final result
[Blue hair]	Gives ‹Blue hair› less weight in the final result
(blue hair:1.4)	Increases emphasis on ‹blue hair› by ~%40

(blue hair:0.6)	Decreases emphasis on ‹blue hair› by ~%40
-----------------	---

Source



These tables demonstrate how parentheses and brackets affect words within prompts for stable diffusion.

In the first table, using parentheses increases emphasis on the enclosed word, with the emphasis factor compounding as the number of parentheses increases. For example, (n) represents a factor of 1.1, while (((n))) represents a factor of 1.331.

In the second table, brackets are used to decrease emphasis, with the reduction becoming more pronounced as the number of brackets increases. For instance, [n] corresponds to a factor of 0.909, while [[[[[n]]]]] represents a factor of 0.564.

Syntax	Equivalent
(n)	(n:1.1)
((n))	(n:1.21)
(((n)))	(n:1.331)
((((n))))	(n:1.4641)
(((n)))	(n:1.61051)
(((n)))	(n:1.771561)

Syntax	Equivalent
[n]	(n:0.9090909090909091)
[[n]]	(n:0.8264462809917355)
[[[n]]]	(n:0.7513148009015778)
[[[[n]]]]	(n:0.6830134553650707)
[[[[[n]]]]]	(n:0.6209213230591552)
[[[[[[n]]]]]]	(n:0.5644739300537775)

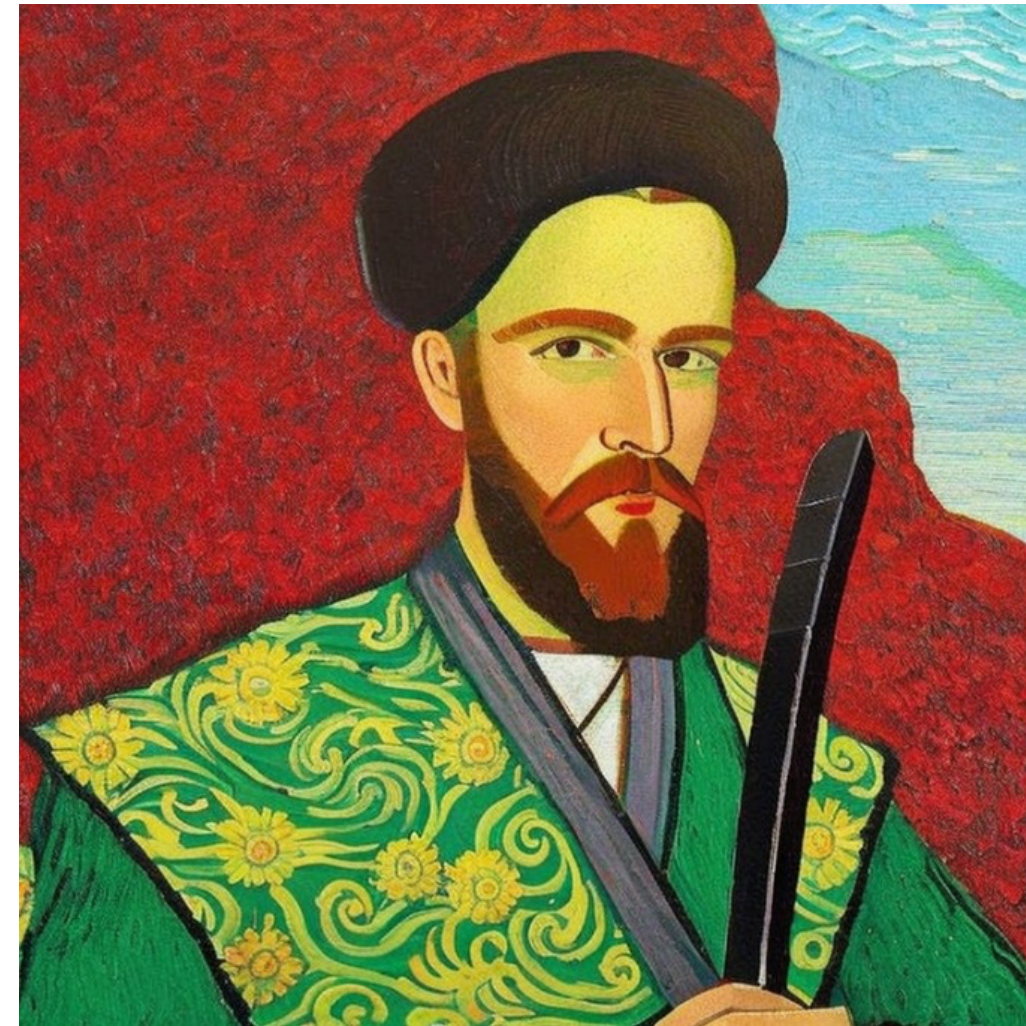


The (word:weight) Method

Utilizing this method involves appending a colon and a decimal number to the keyword you want to assign a weight to. This decimal figure signifies the weight or importance of that keyword, with larger numbers indicating greater importance.

Take this prompt for instance: (Dog:2.0), cute. This instruction suggests to the model that the keyword 'dog' is doubly significant as compared to the keyword 'cute', thereby instructing it to generate an image of a cat with a stronger emphasis.

Method 1: Weight Examples



Prompt: A portrait of a young man standing on a green mountain, holding a sword, in the style of Van Gogh

With no emphasis in the prompt on specific keywords, the image generated was missing multiple factors that I mentioned in the prompt.



Prompt: A portrait, of a (young man:1.5), (standing on a green mountain:1.5), (holding a sword:1.8), in the style of Van Gogh

By adding prompt weights, I was able to get a full standing figure of the man, standing on a green mountain and holding a sword, and although the term "portrait" most often does not result in full-body frames, it did generate one because of the weight placed on "standing".

The Parentheses & Square Brackets Method

To utilize this method, you simply encapsulate the keyword you wish to weight within parentheses. For heightening the keyword's significance further, add more layers of parentheses.

Observe these three examples, where the first word carries the least weight while the fourth word carries the most weight:

A young (boy), ((standing)), (((on a rooftop)))

Should you wish to lessen the importance of a keyword, incorporate it within a square bracket. For instance:

A young [boy], (standing), on a rooftop



Method 2: Weight Examples



Prompt: A portrait of a young man standing on a (green) mountain, holding a sword, in the style of Van Gogh

By emphasizing the word “green” rather than the entire line “standing on a green mountain” the model gave prioritization to the color green without assigning it to a specific keyword in the prompt. And the sword was not taken into enough consideration as an element of the artwork.



Prompt: A portrait, of a ((young)) man, (standing, on a green) mountain, holding a sword, in the style of Van Gogh

By assigning the weights here, I was able to ensure that the man’s clothing was not green, and make the man’s facial features look younger.



Prompt: A portrait, of a ((young)) man, (standing on a green mountain), (((holding a sword))), in the style of Van Gogh

Now none of these 3 examples is a full enough or comprehensive or perfect prompt, but you can clearly see how the parantheses affected the final result.

Sampling



Sampling in Stable Diffusion involves generating new images from a high-dimensional 'latent space'. This process is carried out by repeatedly adding controlled noise to a random image, making it increasingly realistic over time. The frequency of noise application, called 'sampling steps', can be adjusted to balance image quality and generation time. This versatile tool can generate a wide range of images, from realistic and abstract to those that challenge real-world constraints.

A Technical Explanation of Sampling



Sampling, within the context of Stable Diffusion, involves creating novel images from a latent space, a high-dimensional realm comprising all potential images. Stable Diffusion, a form of machine learning model, uses this latent space to fabricate images.

The sampling operation in Stable Diffusion entails the systematic addition of noise to a random image. This noise is incorporated in a controlled manner, enabling the image to incrementally gain realism. The number of noise applications is termed “sampling steps.”

The quantity of sampling steps can be altered to regulate the quality of the ensuing image. An increased count of sampling steps often equates to a more realistic image, but also demands a more extended generation duration.

In order to better understand Stable Diffusion's process with its sampling steps, let's walk through its operation:

Stable Diffusion starts the process by generating a random image within the latent space. This initial image doesn't resemble any real-life depiction.

Next, it introduces noise to this random image. The intensity of the added noise is meticulously regulated according to the number of sampling steps.

With the controlled addition of noise, the image gradually starts to take on a more realistic form.

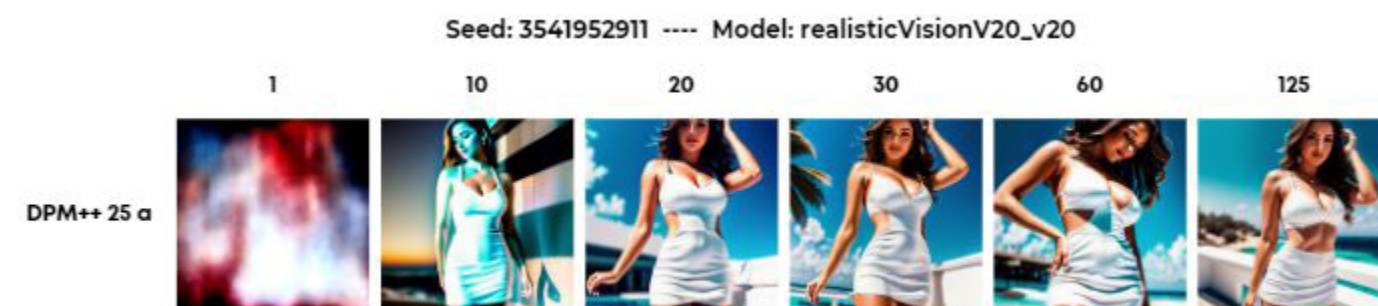
This process of adding noise and subsequently denoising the image is repeatedly carried out until the prescribed number of sampling steps is reached.

The final outcome of this methodical process is an image that represents the culmination of the sampling journey, delivering a product that initially started from pure randomness.

Despite its complexity, the sampling process in Stable Diffusion proves to be a powerful tool for generating a wide variety of images.

Sampling Steps

Sampling steps in Stable Diffusion relate to the frequency of noise application to a randomly generated image within a high-dimensional latent space. This noise is added systematically to make the image progressively more realistic. Adjusting the number of these steps can help balance image quality and generation time. Essentially, a higher number of sampling steps usually equates to a more realistic image, but also requires a longer generation process



Sampling Methods

Stable Diffusion employs an array of sampling techniques to produce images, each contributing to the diversity of the model's output. The chosen sampling method can significantly influence the final image in terms of its overall quality, style, level of detail, and even how closely it aligns with the initial prompt.

For instance, certain sampling methods may better preserve intricate details, while others might lend themselves to producing images with a more abstract or stylized flair. Additionally, some methods might be more apt at creating realistic portrayals, whereas others may excel in bringing to life more fantastical or outlandish concepts.



Sampling Methods

Stable Diffusion utilizes a variety of commonly used sampling methods, including:

DDIM (Denoising Diffusion Implicit Model): A widely favored method for stable diffusion, DDIM is a quick and efficient sampling method capable of producing high-quality images. It is especially adept at creating realistic, smoothly rendered, and painterly-style images.

Euler a: This is a more recent sampling method still being developed. Although more computationally intensive than DDIM, Euler a is capable of generating images with superior detail and a more realistic style.



Sampling Methods

The choice of sampling method is influenced by multiple factors, such as the desired image quality, available time, and computational resources at hand. Here are some considerations when choosing a sampling method:

If speed and efficiency are paramount, DDIM serves as an excellent choice.

For detailed and realistic-style images, Euler a is recommended.

If you're constrained by computational resources, consider a less computationally demanding method, like k-LMS.

In short, the chosen sampling method can substantially influence both the quality and style of the produced image. Gaining an understanding of the diverse sampling methods, along with their respective strengths and limitations, aids in selecting the optimal method for your specific needs.





Rundown of the Various Image Sampling Methods:

DDIM: Also known as the Denoising Diffusion Implicit Model, DDIM is renowned for its quick and efficient generation of high-quality images.

Euler a: Though it demands more computing power than DDIM, Euler a, still in its development phase, promises images with greater detail.

PLMS: Short for Pseudo Linear Multi-Step method, PLMS is a tad faster than DDIM but generates images with slightly less detail.

k-LMS: The Kernelized Linear Multi-Step method is less resource-intensive than DDIM and PLMS. While its image detail is lower, it's a viable choice if you're resource-limited.

Fast DDIM: A speedier version of DDIM under development, Fast DDIM, trades a little accuracy for speed but remains an efficient choice.

Adaptive DPM: Derived from the Denoising Diffusion Prior Model, Adaptive DPM tailors its process to the image content, leading to more realistic outcomes.

Euler b: Still under development, Euler b demands more computational resources than Euler a, but produces more detailed and realistic images.

DDIM-GAN: This hybrid model combines DDIM with a Generative Adversarial Network (GAN) to deliver higher detail and realism than DDIM alone.

Now, let's dive deeper into a detailed comparison of these sampling methods, examining their respective advantages and disadvantages.

Method	Description	Strengths	Weaknesses
Euler a	Simple, predictable sampling method without random noise.	Generates smooth, detailed images with good color blending.	Can be slow and occasionally produces imperfect artifacts.
Euler	Advanced sampling method using a diffusion model to manage image noise.	Yields sharper images with less noise than Euler a.	Can be slower than Euler a and may have less detail.
LMS	Utilizes a low-memory variant of the Karras diffusion model.	Produces fast, high-quality images with good detail.	Can produce images with less detail compared to others.
Heun	Incorporates a Heun integrator to improve sampling stability.	Produces smoother images than Euler a or LMS.	May be slower than Euler a or LMS.
DMP2	Uses a diffusion model with larger capacity than the original DPM model.	Produces high-quality images with good detail.	Can be slow and requires more training data than others.
DMPa	Utilizes a diffusion model with smaller capacity than the original DPM model.	Generates faster images with less detail than DMP2.	May produce images with less detail and color blending.

Now, let's dive deeper into a detailed comparison of these sampling methods, examining their respective advantages and disadvantages.

Method	Description	Strengths	Weaknesses
DDIM	Uses a diffusion model to predict image noise and color, resulting in high-quality realism.	Produces high-quality, realistic images with good detail.	Slower and requires a larger training dataset.
PLMS	Applies a progressive learning method to improve image quality over time.	Produces high-quality images with good detail.	Can be slow and requires more training data than others.
LMS Karras	Utilizes a memory-efficient version of the Karras diffusion model.	Produces fast, high-quality images with good blending.	May have slightly less detail than other samplers.
DPM Adaptive	Uses a diffusion model with larger capacity than the original DPM model.	Produces high-quality images with good detail.	Can be slow and requires more training data than others.
DPM2 Karras	Uses a diffusion model with larger capacity than the original DPM model.	Produces high-quality images with good detail.	Can be slow and requires more training data than others.
DMPa	Combines larger-capacity diffusion model with memory-efficient Karras model.	Generates fast, high-quality images with good blending.	May have slightly less detail than other samplers.

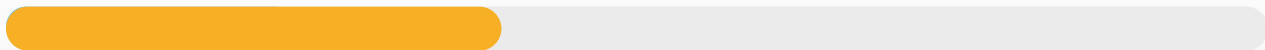
Now, let's dive deeper into a detailed comparison of these sampling methods, examining their respective advantages and disadvantages.

Method	Description	Strengths	Weaknesses
DPM2 a Karras	Uses a smaller-capacity diffusion model, low-memory Karras model, produces faster images with less detail	Fast, high-quality images with good detail, color blending	Less detail and color blending compared to DPM2 Karras
DPM++ 2M Karras	Utilizes a larger-capacity diffusion model, 2 million step sampling schedule	High-quality images with good detail, color blending	Slow, requires more training data
DPM++ 25 a	Employs a larger-capacity diffusion model, -25step sampling schedule	High-quality images with good detail, color blending	Slow, requires more training data
DPM++ 2M	Utilizes a larger-capacity diffusion model, 2 million step sampling schedule	High-quality images with good detail, color blending	Slow, requires more training data
DPM++ SDE	Uses a larger-capacity diffusion model, stochastic differential equation sampling schedule	High-quality images with good detail, color blending	Slow, requires more training data
DPM Fast	Utilizes a smaller-capacity diffusion model, fast sampling schedule	Fast images with less detail compared to DPM2	Less detail and color blending compared to DPM2



Now, let's dive deeper into a detailed comparison of these sampling methods, examining their respective advantages and disadvantages.

Method	Description	Strengths	Weaknesses
DPM++ SDE Karras	Utilizes a larger-capacity diffusion model, stochastic differential equation sampling schedule	High-quality images with good detail, color blending	Less detail and color blending compared to DPM2 Karras
PLMS	Uses a progressive learning method to improve image quality	High-quality images with good detail, color blending	Slow, requires more training data
UniPC	Utilizes a unified diffusion model to predict image noise and color	High-quality images with good detail, color blending, realism	Slow, requires more training data

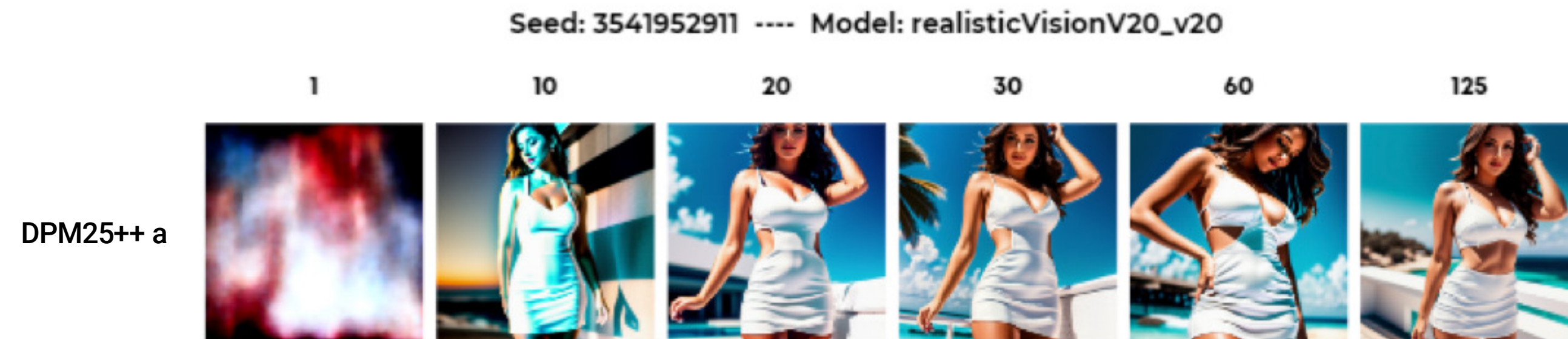


This comparative chart offers a transparent visual illustration of the impact that different sampling methods and sampling steps have on the resulting image.

Each image represented in the chart has been generated utilizing an identical prompt, seed, and Configuration File (CFG) parameters. The only variables altered in this experiment are the sampling steps (which are represented on the x-axis), and the sampling methods (shown on the y-axis).



These are additional examples illustrating the same principle, but exclusively employing Euler a, DDIM, and DPM25++ a as the sampling methods. They again underscore the significant role that specific sampling techniques play in the outcome of the image, even when the prompt, seed, and Configuration File (CFG) parameters remain constant.



Seed



A seed is a numerical value vital for initiating image generation. It enables the production of reproducible images and allows for experimentation with parameters or prompt variations, while maintaining consistent outcomes. If you don't specify a seed, one is randomly generated. However, having control over the seed gives you greater influence over the image generation process.

It's important to remember what a seed isn't. While it's a number used to generate noise, it's not the image of noise itself, it doesn't contain all parameters used in image generation, and it's not linked to a specific text prompt.

Every image created has a unique "Seed" attribute, effectively making it the image's identity. This seed can allow anyone to generate the exact same image, with room for variations. So, the seed holds the key to image reproducibility and control over the image generation process.

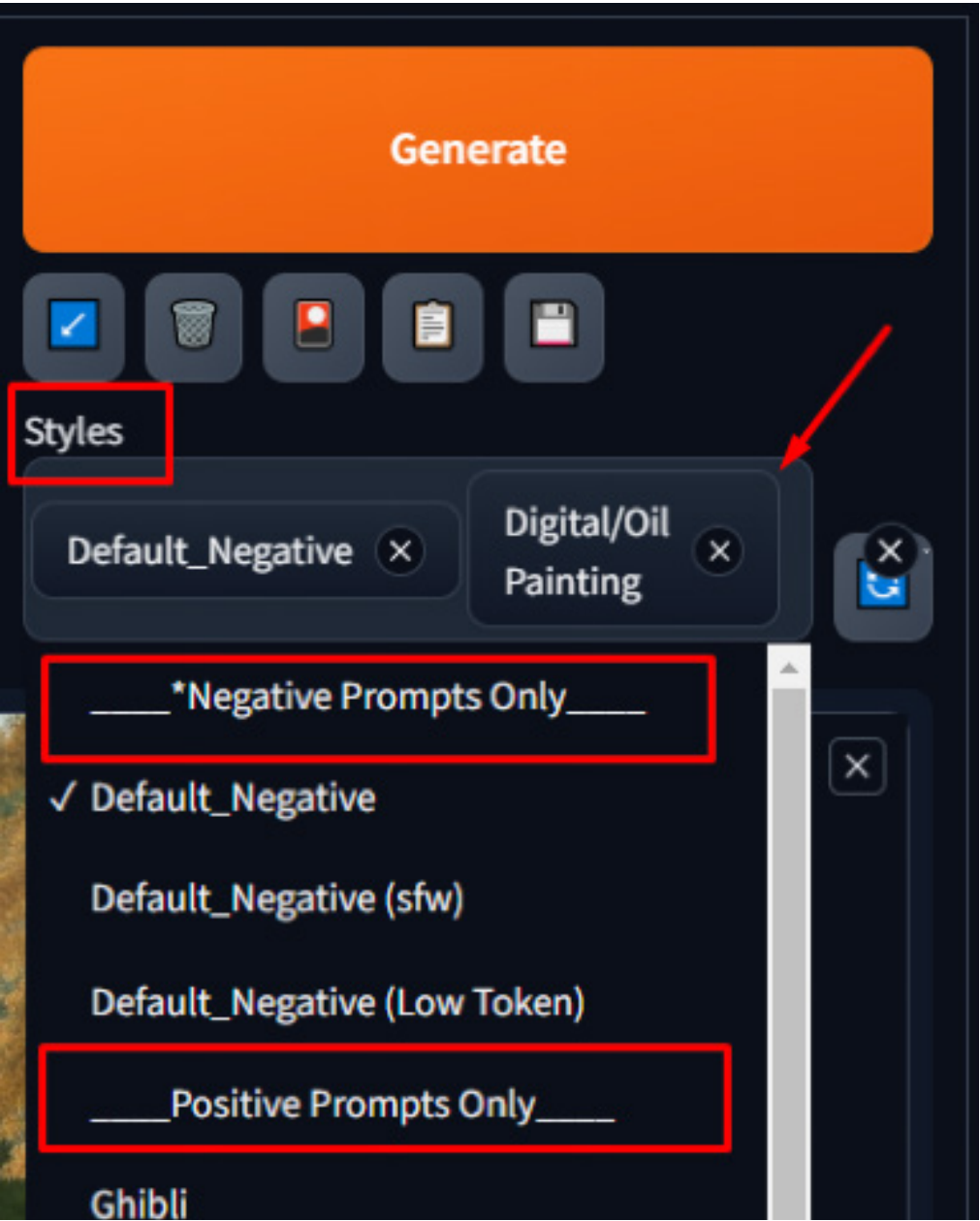
Styles & Presets

A	B	
name	prompt	negative_prompt
____*Negative Prompts Only____		
Default_Negative		canvas frame, (high contrast:1.2)
Default_Negative		NSFW, Cleavage, Pubic Hair, Nu
Default_Negative		lowres, bad anatomy, bad hand
____Positive Prompts Only____		
Ghibli	(Studio ghibli style, Art by Hayao Miyazaki:1.2), An	
Vector Illustratio	Vector art, Vivid colors, Clean lines, Sharp edges, N	
Digital/Oil Painti	(Extremely Detailed Oil Painting:1.2), glow effects,	
Indie Game	Indie game art,({prompt}), (Vector Art, Borderland	
Original Photo S	Photorealistic, Hyperrealistic, Hyperdetailed, analc	
Black and White	(b&w, Monochromatic, Film Photography:1.3), Phc	
Isometric Rooms	Tiny cute isometric {prompt} in a cutaway box, soft	
Space Hologram	hologram of {prompt} floating in space, a vibrant d	
Cute Creatures	3d fluffy {prompt}, closeup cute and adorable, cute	
Realistic Photo P	RAW candid cinema, 16mm, color graded portra 40	
Professional Sce	long shot scenic professional photograph of {prom	
Manga	(Manga Style, Yusuke Murata, Satoshi Kon, Ken Sug	
Anime	(Anime Style, Tezuka Osamu, Satoshi Kon, Ken Sug	

Utilize the "Save Style" feature when you discover an image generated using your ideal keywords. This function allows you to designate distinct positive and negative prompts that can subsequently be employed as "templates" or "presets", a time-efficient strategy commonly used by professionals. Undoubtedly, this will expedite your path towards achieving consistent high-quality results with your prompts.

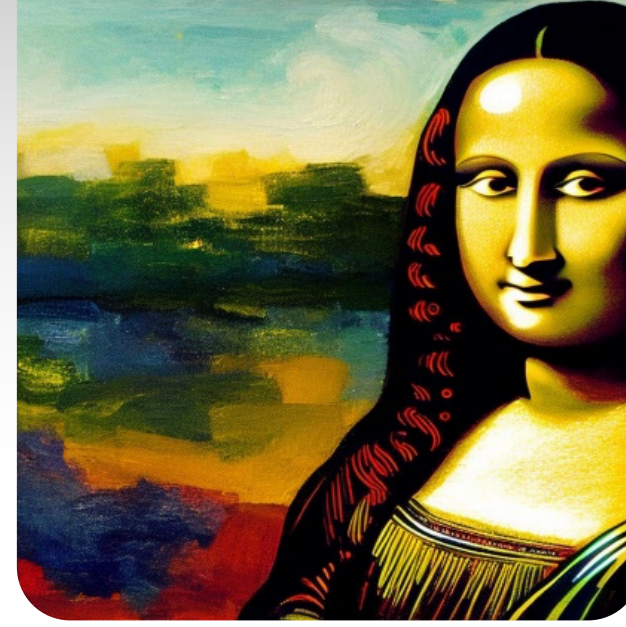
I recommend using prompt templates both for positive and negative prompts and importing them into stable diffusion. This is another way you can save a set of style presets or import in bulk. Create a "Styles" CSV file with three columns: "name", "prompt", and "negative_prompt". Add category titles in the name column, such as "Negative Prompts" and "Positive Prompts". For negative prompts, add a unique name for each set and corresponding prompts in the "negative_prompts" column, leaving the "prompt" column blank. Repeat for positive prompts, but place the prompt in the "prompt" column. Save the file in your webui folder without subfolders. You can then select default styles from your GUI's top-right menu. Professional presets, like [Sebastian Kamph's](#), are also available for download.

Note that you can even download presets that professionals use such as [Sebastian Kamph's](#) [here](#).

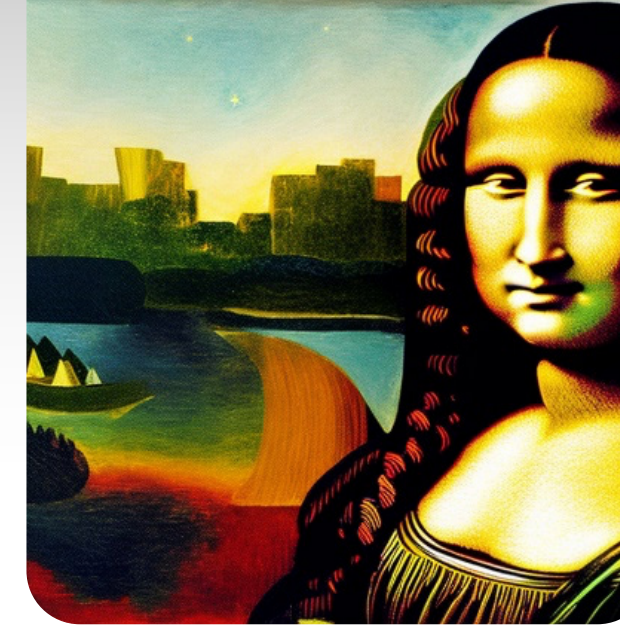




[Mona Lisa: Starry Night: 0.1]



[Mona Lisa: Starry Night: 0.5]



[Mona Lisa: Starry Night: 0.75]

Prompt Scheduling

Prompt scheduling is a technique that allows you to merge two keywords in a particular syntax, essentially enabling you to merge content and images in a single prompt without using img2img prompting: [keyword1 : keyword2: factor].

The 'factor' is a value ranging between 0 and 1 that determines the point at which the transition from keyword1 to keyword2 occurs.

As an illustration, consider the prompt:
"Abstract art representation of [Mona Lisa: Starry Night: 0.4]"

If we take this prompt through 50 sampling steps, the interpretation for the first 20 steps (50 steps x 20 = 0.4 steps) will be:

"Abstract art representation of Mona Lisa"

Then, for the remaining 30 steps, the interpretation shifts to:
"Abstract art representation of Starry Night"

The factor in this case dictates the transition between the two themes. Manipulating this factor allows for varying degrees of blending between the two subjects.

Look at these three images for example, you can clearly see that as we increase the numerical factor variable, you see more of the Mona Lisa and less of Starry Night, keeping in mind that the prompt itself (Abstract art representation of [Mona Lisa: Starry Night: 0.75]) was very simple and by using a more advanced and detailed prompt, the possibilities are endless

Remember that the steps play a factor, so having too few steps would make this a difficult-to-illustrate process.

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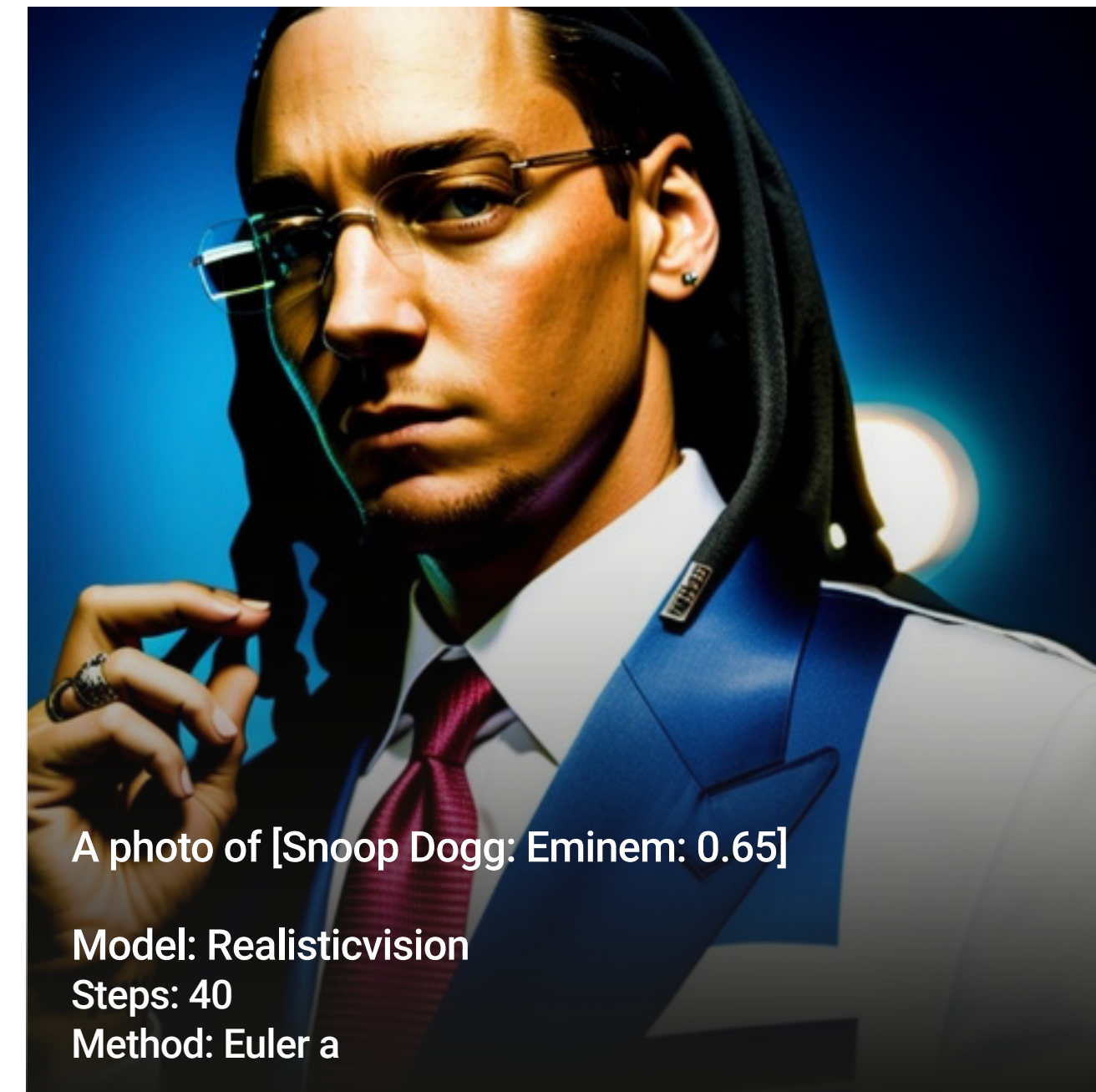
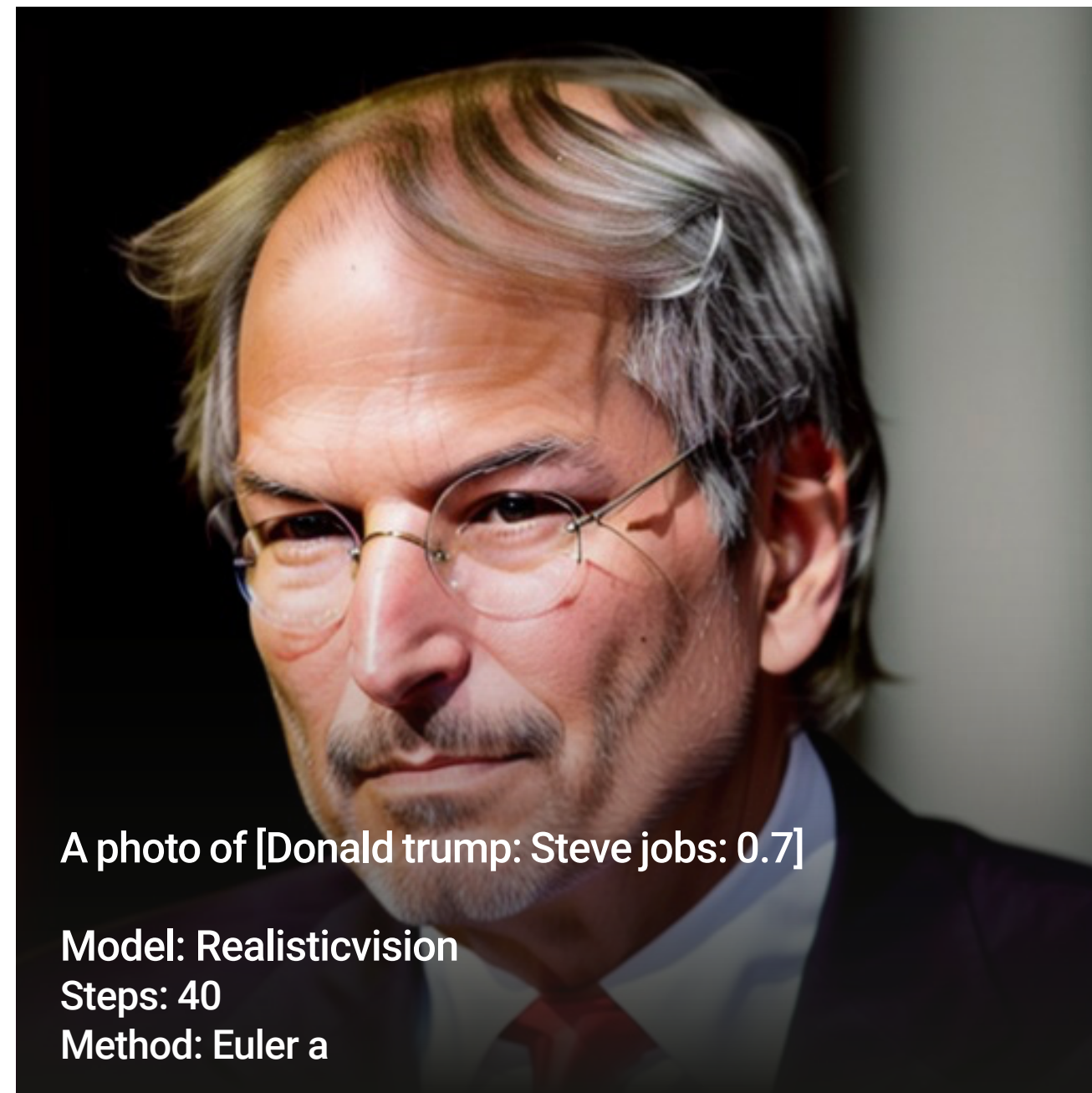
Remember that the steps play a factor, so having too few steps would make this a difficult-to-illustrate process.

Prompt Scheduling

Common Use Case: Face Blending



Fun tip: Pay attention to the process in the GUI, you'll see how throughout the steps, the face-blending changes





CFG Parameter

The CFG scale, or Classifier-Free Guidance scale, is a parameter utilized in Stable Diffusion, notably in Stable Diffusion V2.1. It governs the level of conformity between the generated image and the provided text prompt or input image. This scale determines the degree to which the image generation process aligns with the given input.

So What Does it do Really?

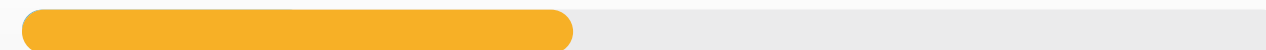
The CFG scale plays a crucial role in adjusting the visual output of the Stable Diffusion model. It serves as a mechanism to balance creativity and adherence to the input prompt or image. By modifying the CFG scale value, you can control the level of conformity between the generated image and the input. Lower values promote greater creativity, enabling the model to deviate more from the prompt, while higher values prioritize closer adherence to the input prompt or image.



Ideal Ways of Using:

The default CFG scale value in certain interfaces, such as the SD GUI, is 7, which strikes a suitable balance between creativity and desired outcomes. However, the optimal CFG scale value can vary depending on the specific use case. For improved results with minimal noise, a CFG scale value between 7 and 11 is generally recommended across the Stable Diffusion GUI and various other models available in the same GUI on Automatic1111.

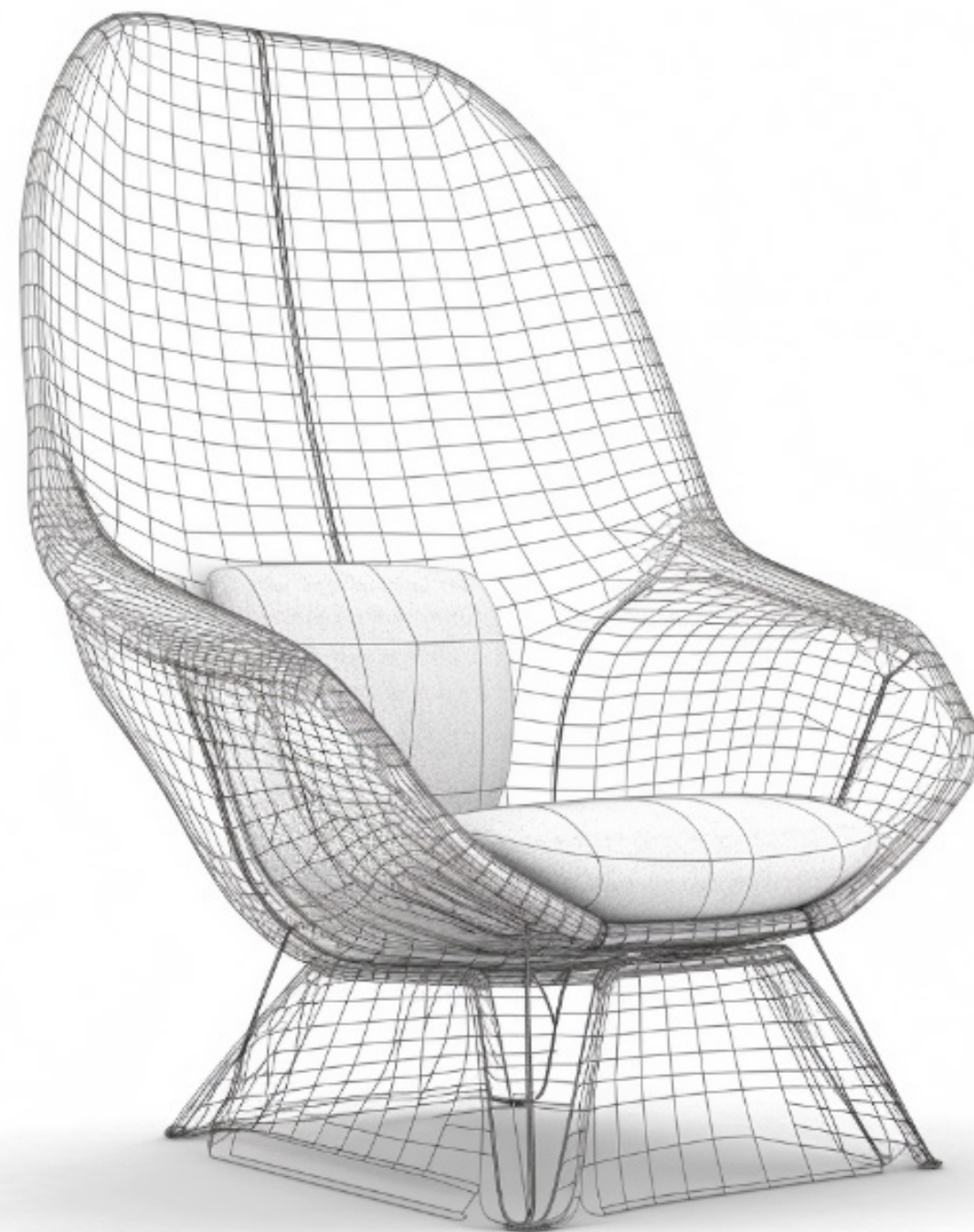
It's essential to find a balance with the CFG scale to avoid excessive complexity in the generated images. Setting the CFG scale too high may result in overcomplicated visuals, as the AI attempts to incorporate every word from the prompt as a detailed element. It's crucial to strike a balance that ensures the desired level of fidelity and coherence in the generated image while avoiding unnecessary complexity.



How it Fits With Everything Else:

The CFG scale operates in tandem with other parameters within Stable Diffusion, particularly complementing the guidance scale. The guidance scale determines the degree of adherence between the generated image and the text prompt, while the CFG scale influences the overall behavior of the model in generating images.

It's worth noting that while Stable Diffusion V2.1 is not explicitly mentioned in the search results provided, the information regarding the CFG scale is generally applicable to the model, including V2.1.



Sampling method: Euler a

Sampling steps: 40

☐ Restore faces ☐ Tiling ☐ Hires. fix

Width: 768

Height: 768

Batch count: 1

Batch size: 1

CFG Scale: 30

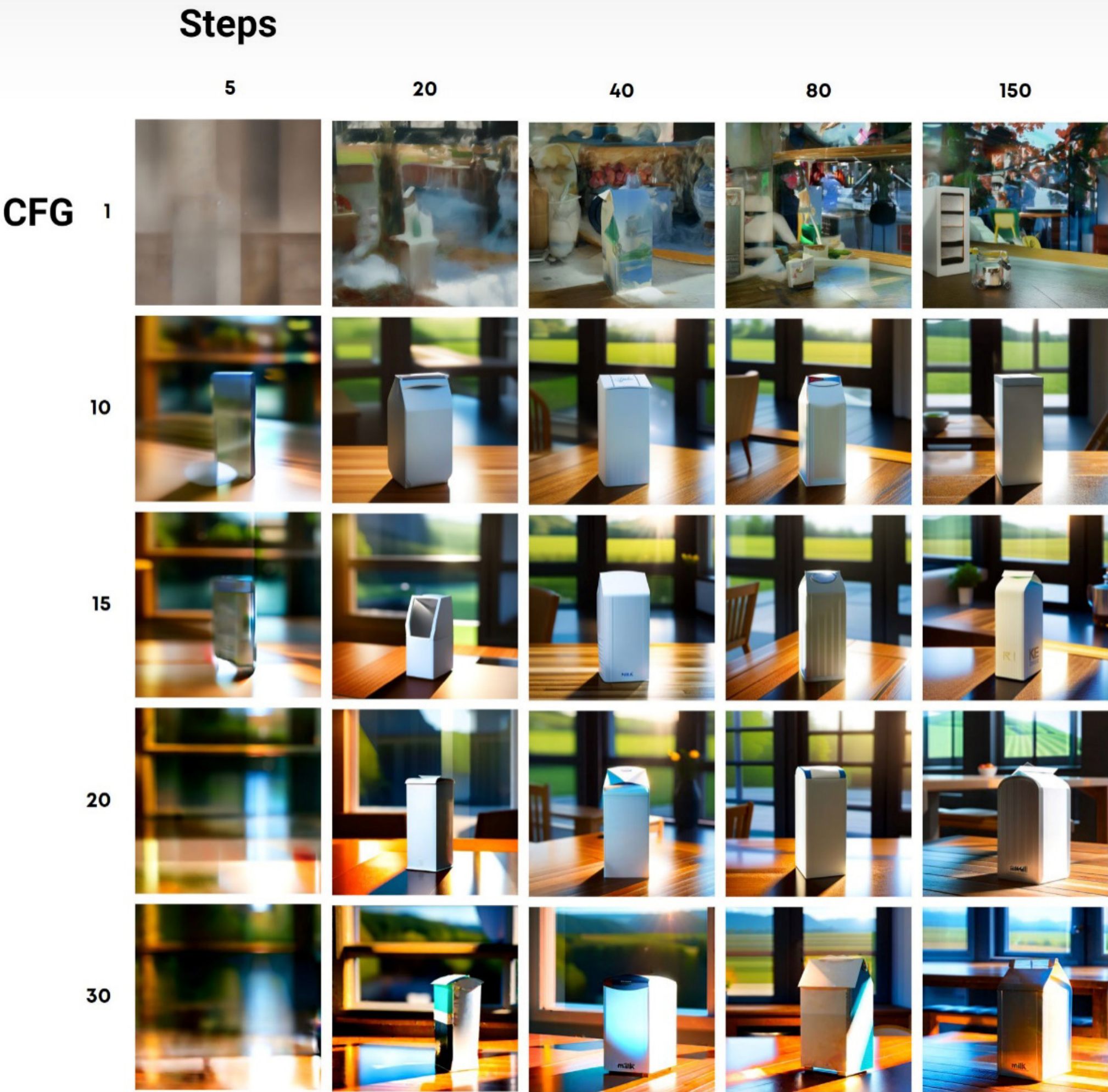
Seed: 2135992340

☐ Extra

Mix and Match

When adjusting the sampling steps and CFG parameter simultaneously in Stable Diffusion, their values interact to shape the resulting image. Increasing the sampling steps enhances detail and complexity, while increasing the CFG parameter reinforces the influence of the text prompt. Higher values for both variables can potentially yield images with refined details, stronger adherence to the prompt, and a closer realization of the desired concept.

However, it's crucial to consider the trade-off. Manipulating these variables together may prolong processing times due to the increased computational resources required. Striking a balance between desired detail and adherence to the prompt, while considering practical constraints of time and computational resources, is key!



Advanced Prompting Examples



Steps: 40
Sampler: Euler a
CFG scale: 7
Seed: 4208389335
Size: 512x512
Model hash: f69b6b6348
Model: deliberateRealistic_v10

Prompt: Iron Man, (Arnold Tsang:Toru Nakayama:0.6), (Masterpiece:1.4), Studio Quality, 1boy, (glowing:1.2), axe, mecha, science_fiction, (solo:1.4), (weapon:1.2), (jungle:1.4), green_background, nature, outdoors, solo, tree, weapon, mask, dynamic lighting, detailed shading, (digital texture painting:1.2), (Extremely Detailed Oil Painting:1.2), glow effects, godrays, render, 8k, octane render, cinema 4d, blender, dark, atmospheric 4k ultra detailed, cinematic sensual, Sharp focus, humorous illustration, big depth of field, Masterpiece, colors, 3d octane render, 4k, concept art, trending on artstation, hyperrealistic, Vivid colors, extremely detailed CG unity 8k wallpaper, trending on ArtStation, trending on CGSociety, Intricate, High Detail, dramatic, absurdes

Negative prompt: un-detailed skin, semi-realistic, cgi, 3d, render, sketch, cartoon, drawing, ugly eyes, (out of frame:1.3), worst quality, low quality, jpeg artifacts, cgi, sketch, cartoon, drawing, (out of frame:1.1), canvas frame, (high contrast:1.2), (over saturated:1.2), (glossy:1.1), cartoon, 3d, ((disfigured)), ((bad art)), ((b&w)), blurry, ((bad anatomy)), (((bad proportions))), ((extra limbs)), cloned face, (((disfigured))), extra limbs, (bad anatomy), gross proportions, (malformed limbs), ((missing arms)), ((missing legs)), (((extra arms))), (((extra legs))), mutated hands, (fused fingers), (too many fingers), (((long neck))), Photoshop, video game, ugly, tiling, poorly drawn hands, 3d render



Steps: 40
Sampler: DDIM
CFG scale: 9
Seed: 2570480927
Size: 512x512
Model hash: f69b6b6348
Model: deliberateRealistic_v10

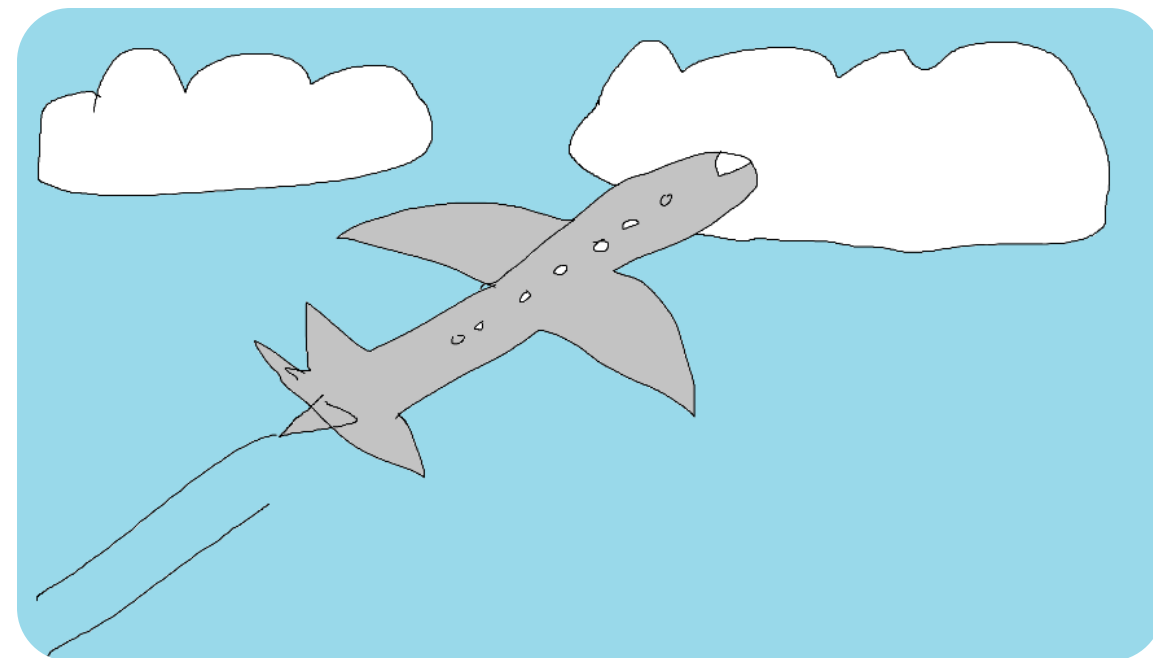
Prompt: (ancient scroll vector design:1.2), (bold shūji:1.1), chart, schematics, infographic, scientific, measurements, abstract, surreal, collage, new media design, poster, colorful highlights, tarot card, glowing ruins, marginalia, 8k, (extremely detailed:1.2), (style of Katsuhiko Otomo and Masamune Shirow:1.4) pantone, on black canvas, (typography annotations:1.3), Vector art, Vivid colors, Clean lines, Sharp edges, Minimalist, Precise geometry, Simplistic, Smooth curves, Bold outlines, Crisp shapes, Flat colors, Illustration art piece, High contrast shadows, Technical illustration, Graphic design, Vector graphics, High contrast, Precision artwork, Linear compositions, Scalable artwork, Digital art

Negative prompt: un-detailed skin, semi-realistic, cgi, 3d, render, sketch, cartoon, drawing, ugly eyes, (out of frame:1.3), worst quality, low quality, jpeg artifacts, cgi, sketch, cartoon, drawing, (out of frame:1.1), canvas frame, (high contrast:1.2), (over saturated:1.2), (glossy:1.1), cartoon, 3d, ((disfigured)), ((bad art)), ((b&w)), blurry, ((bad anatomy)), (((bad proportions))), ((extra limbs)), cloned face, (((disfigured))), extra limbs, (bad anatomy), gross proportions, (malformed limbs), ((missing arms)), ((missing legs)), (((extra arms))), (((extra legs))), mutated hands, (fused fingers), (too many fingers), (((long neck))), Photoshop, video game, ugly, tiling, poorly drawn hands, 3d render

img2img Prompting

The img2img prompting process within Stable Diffusion begins with the usage of a Large Language Model (LLM) that formulates a detailed textual depiction of the desired image. This written narrative guides the Stable Diffusion model in creating the corresponding image.

Initially, the Stable Diffusion model conjures up a random image. As the process progresses, this image is iteratively fine-tuned, gradually growing closer to the envisioned depiction from the LLM. This refining stage involves a controlled procedure of noise introduction, enhancing the image details in a stepwise manner, causing the image to gradually align with the desired concept. In essence, the text narrative is effectively translated into a tangible visual manifestation.



A sketch drawn on paint in 30 seconds



A very basic prompt in the img2img tab turned into this nearly-decent photo



Possibilities

Img2img prompting in Stable Diffusion unveils a vast landscape of possibilities, permitting the creation of a diverse range of imagery:

Realistic Illustrations: By harnessing img2img prompting, you can create images exhibiting a high degree of realism, comparable to actual photographic captures, permitting the depiction of tangible scenes and objects.

Inventive Visuals: Not just confined to realistic representation, img2img prompting paves the way for generating creative, even otherworldly imagery that exceeds the capabilities of standard photography.

Tailored Imagery: Moreover, img2img prompting can craft custom-made visuals catering to personal tastes or explicit needs, fashioning images particularly attuned to one's preferences or requirements.

Consequently, img2img prompting within Stable Diffusion spans a broad spectrum of image generation scenarios.

And yes, it even allows you to transform yourself into a Disney character!

Img2img Value

Img2img prompting in Stable Diffusion is a crucial asset for creative endeavors. It's a tool for generating diverse images that can be utilized in:

Art: Artists can employ this technology to create diverse forms of art, enriching their creative expressions.

Design: It aids in visualizing product designs, streamlining the creation of anything from furniture to clothing.

Entertainment: In films, TV shows, and video games, img2img prompting assists in crafting captivating visuals and narratives.

Sketch to Photo

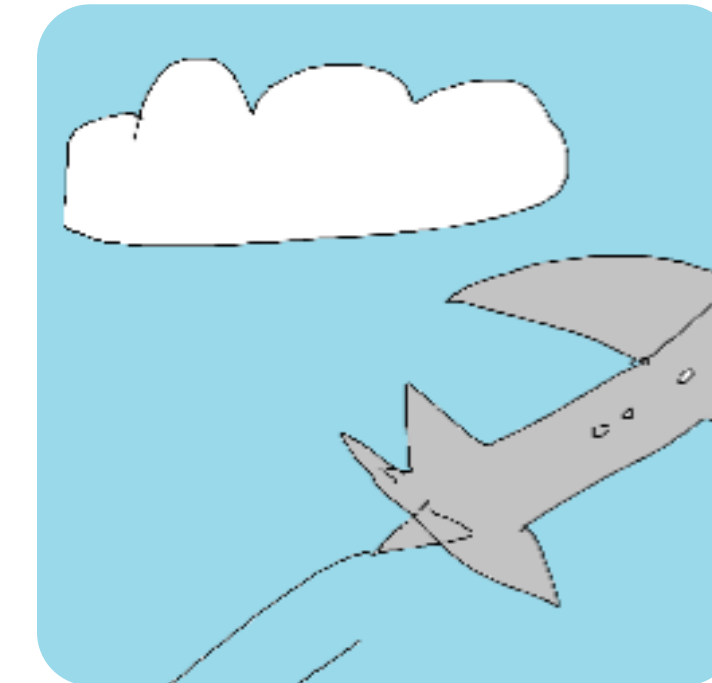
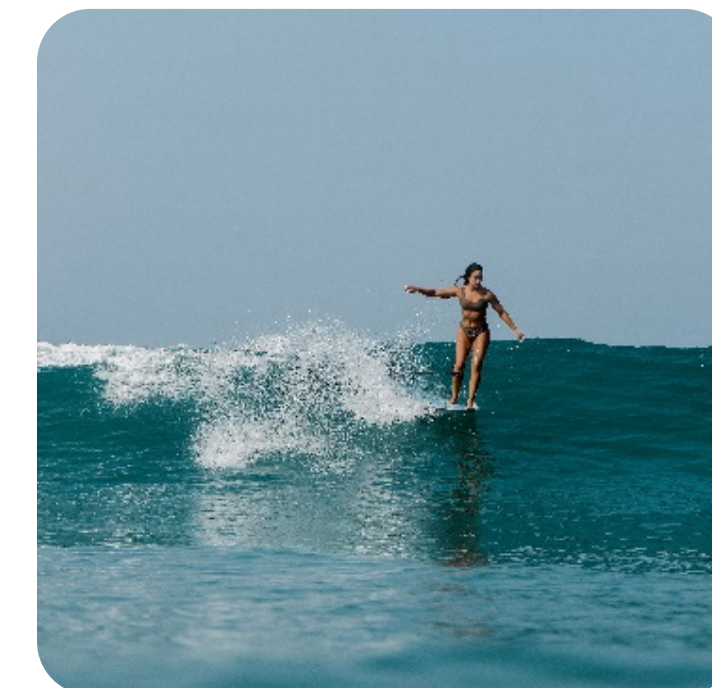


Photo to Painting



img2img Prompting:

While these suggestions are more general in nature, the exact number of steps necessary might differ based on the specific image you aim to create. It could take some trial and error with different step counts to discover what fits your needs best.

Bear in mind that increasing the steps also elongates the image generation time. Thus, if you're constrained by time, it might be beneficial to opt for fewer steps. However, it's worth noting that a lesser number of steps might compromise the quality of the final image.

Here are some further pointers when working with Stable Diffusion sampling methods:

Opt for high-quality prompt images. The higher the prompt image's quality, the better the resultant generated image will be.

Precision is key in your prompts. The more accurate your description, the higher the likelihood that the output image aligns with your anticipation.

Utilize a progressive learning approach. Such methods can incrementally enhance the generated image's quality.

Experiment with diverse sampling techniques. There's an array of available sampling methods, each with its unique advantages and downsides. It's beneficial to explore these various methods to identify the one that fits your needs best.

Sampling Method	Recommended Sampling Steps
Euler a	30-20 steps
DDIM	40-30 steps
DPM++ 25a	50-40 steps

Denoising Strength

In Stable Diffusion's img2img application, 'denoising strength' represents the level of noise removal applied during image creation. This value ranges from 0 (no denoising) to 1 (maximum noise removal).

In the simplest terms, you can control the extent of detail using this option, basically having a lower denoising strength will result in less detailed interference in this process, while a higher denoising strength will have a more detailed effect, meaning if you want to use this to fix facial expressions for example, you'd want to go higher on denoising strength.

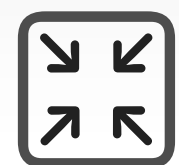
The ideal denoising strength relies on various factors:

- Noise Level: More denoising can help eliminate major noise or unwanted elements, yielding a smoother result.
- Detail Preservation: Too much denoising might erase subtle details. If you want to keep textures and details, use less denoising.
- Artistic Goals: If maintaining specific details is important for your artwork, you might want to use less denoising.
- Trial and Error: Gradually tweaking the denoising strength from a middle value can help strike a balance between noise removal and detail preservation.
- Personal Preference: Ultimately, the best denoising strength depends on your personal taste and the requirements of your project. Don't hesitate to experiment to find what works best for you.



The screenshot shows the Stable Diffusion web interface with various settings. The 'Denoising strength' slider is highlighted with a red box and is set to 0.75. Other visible settings include Width (512), Height (512), CFG Scale (7), Batch count (1), and Batch size (1). The Seed is set to -1, and the ControlNet version is v1.1.215. The Regional Prompter is set to None.

Parameter	Value
Width	512
Height	512
CFG Scale	7
Denoising strength	0.75
Seed	-1
ControlNet v1.1.215	Regional Prompter
Script	None



In Painting

Inpainting is a technique used to fill in (or replace) missing or damaged portions of an image.

Inpainting is a technique used to fill in (or replace) missing or damaged portions of an image. It's frequently employed in restoring damaged photographs or creating new images from the ground up. While this process can be executed manually, it can also be efficiently automated using computer algorithms.

There are several potential applications of inpainting, which include:

- **Repairing Damaged Images:** Inpainting is an effective tool for fixing damaged photographs, for instance, those that are scratched, torn, or faded. Additionally, it aids in removing unwanted elements from images, like power lines or people.
- **Creating New Images:** Inpainting is capable of generating new images from scratch, offering utility in producing creative content or exploring novel artistic avenues.

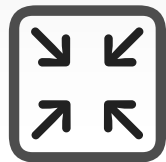
Despite its potential, there are certain challenges associated with inpainting:

Difficulty in Control: It can be challenging to steer the model to generate the exact image desired. As the model is trained on a vast dataset of images, controlling the specific output can be complex.

Time-Consuming: Generating a high-quality image using inpainting can be a time-intensive process.

Computationally Expensive: Running the diffusion model for inpainting necessitates a powerful computer, making it a computationally demanding process.





In Painting

First, let's introduce a helpful trick

If you have previously worked with inpainting, you may have faced the challenge of zooming into the canvas, given that zoom functionality is not supported by default in the Automatic1111 GUI. However, there is a workaround!

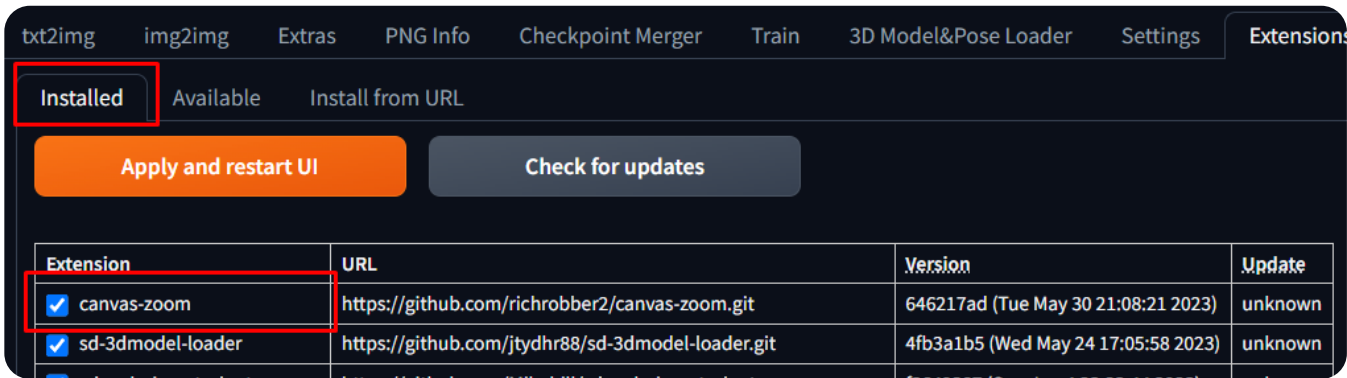
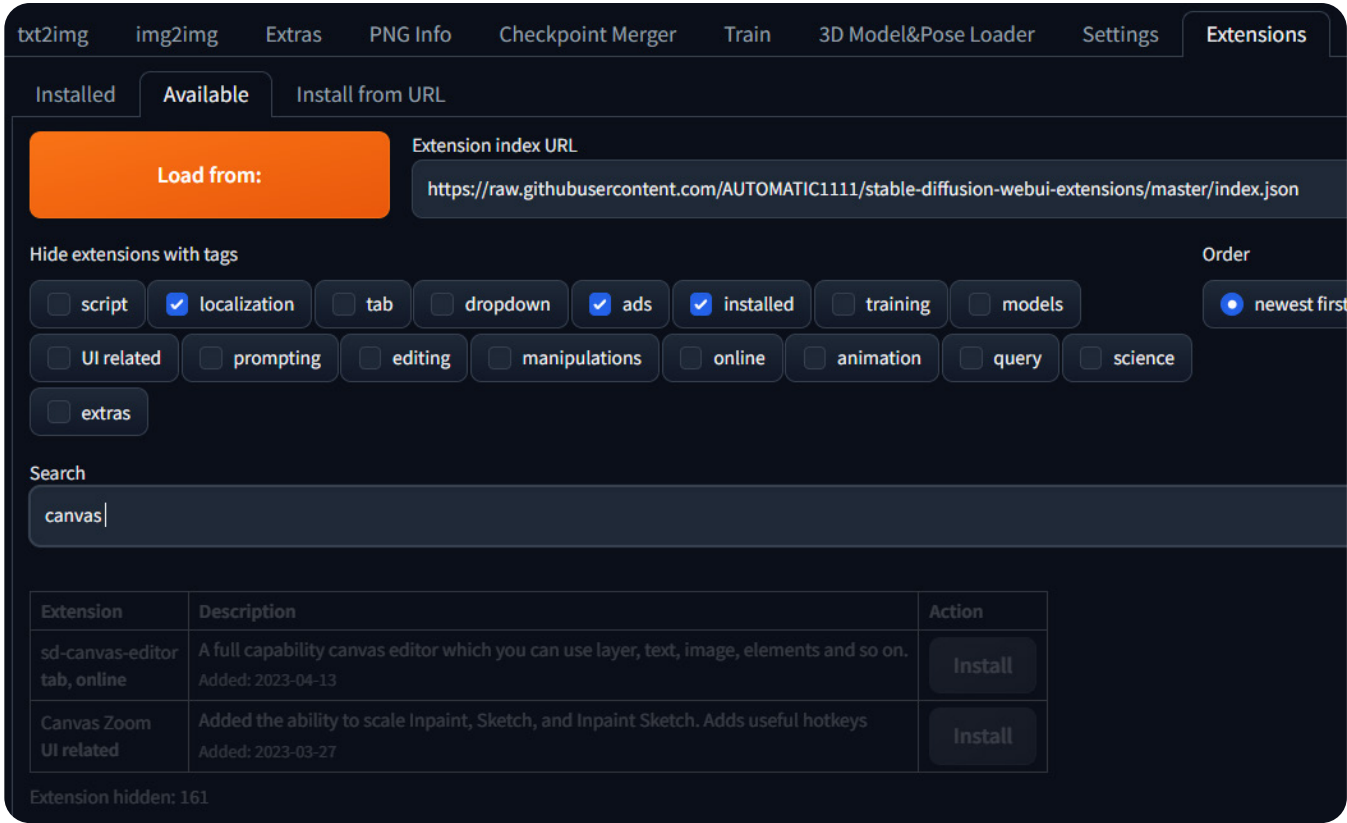
Navigate to the extensions tab and ensure that you're on the "Available" sub-tab. Input the following link into your "Extension index URL" tab:

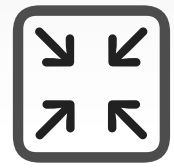
<https://raw.githubusercontent.com/AUTOMATIC1111/stable-diffusion-webui-extensions/master/index.json>

Next, conduct a search for "Canvas" and install the extension named "Canvas Zoom".

To confirm its installation, move to the "Installed" sub-tab and verify its presence in the list of installed extensions. Click on "Apply and restart UI", and there you have it!

Now, you can effortlessly utilize the default Ctrl+scroll for adjusting the brush size and Shift+scroll for zooming in and out of the canvas. This will significantly enhance your comfort and precision while working.





In Painting

Here's a quick overview of the key parameters to keep in mind:

The screenshot displays the Inpainting interface with the following settings:

- Mask mode:** ☒ Inpaint masked, ☐ Inpaint not masked
- Masked content:** ☐ fill, ☒ original, ☐ latent noise, ☐ latent nothing
- Inpaint area:** ☒ Whole picture, ☐ Only masked. The "Only masked padding, pixels" slider is set to 32.
- Sampling method:** Euler a
- Sampling steps:** 20
- Restore faces:** ☐ **Tiling:** ☐

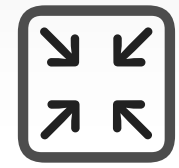
Firstly, consider the "Mask mode". It's quite straightforward – if "Inpaint masked" is selected, the Stable Diffusion (SD) model will apply your prompt to the area you've masked. Conversely, if "Inpaint not masked" is chosen, the SD model will apply modifications to the rest of the image while disregarding the masked area. Each option has its specific applications.

Next, pay heed to the "Masked content" section. If "Original" is selected, as in my current setting, the model will factor in the underlying content of the masked area when producing the new result. This is particularly beneficial when making subtle adjustments to a human face in Inpaint.

If you aim to introduce a new element into an image that didn't exist before, you should select "latent noise".

The "Inpaint area" parameter lets you choose whether to render the entire image or just the masked area. In most instances, it's advisable to opt for "Whole picture", ensuring equal render quality and resolution across the entire image, both within and outside the masked area. This helps maintain consistency in your final result.

Should you select "only masked", the model will render just the masked area and possibly an adjacent region, the size of which you can modify using the "Only masked padding pixels" slider situated next to the "Inpaint area".

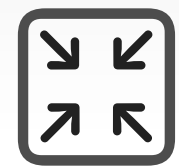


In Painting



Photos by Jeet Dhanoa on Unsplash





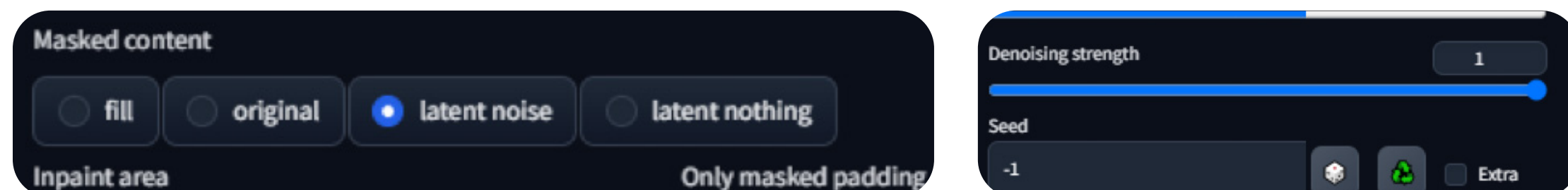
In Painting

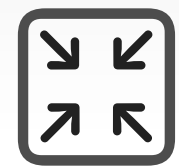
Adding non-existing objects



Photos by Jeet Dhanoa on [Unsplash](#)

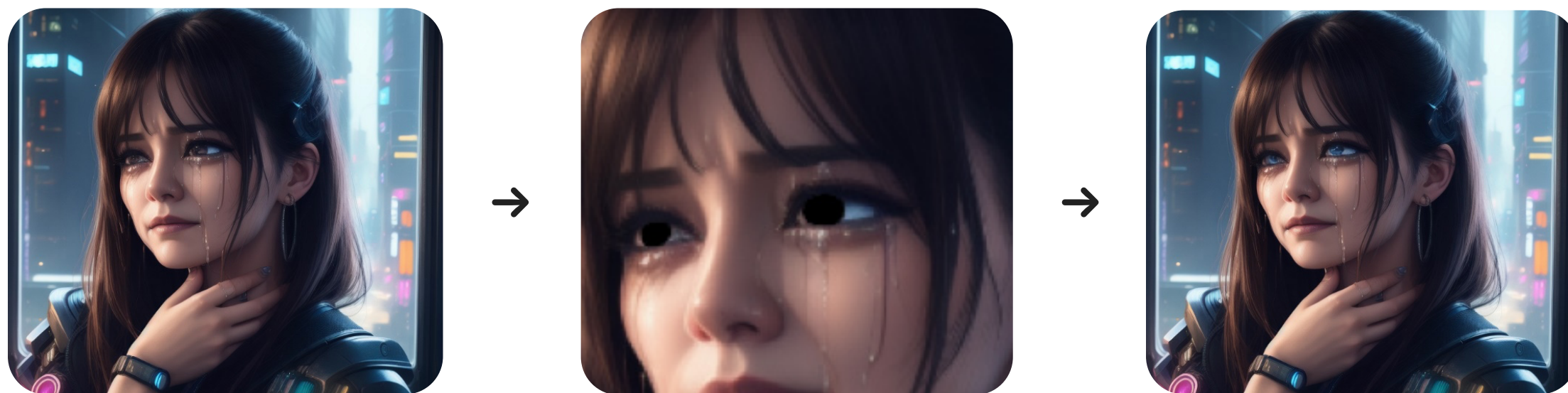
When attempting to add objects that did not exist at all in the original image (as opposed to changing an existing object), it's crucial to select "latent noise", ensuring the model understands that it will not reference the content beneath the masked area. Additionally, boost the denoising strength to its maximum capacity, as a robust detail-creation effort is required for the subsequent render.



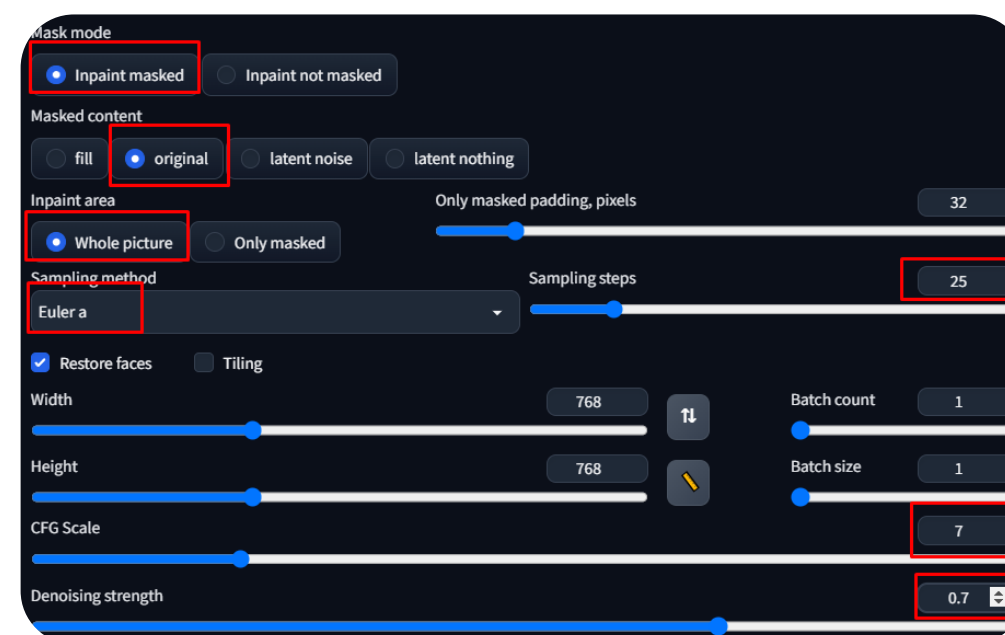


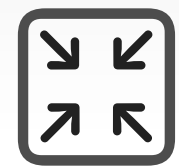
In Painting

Changing eye color



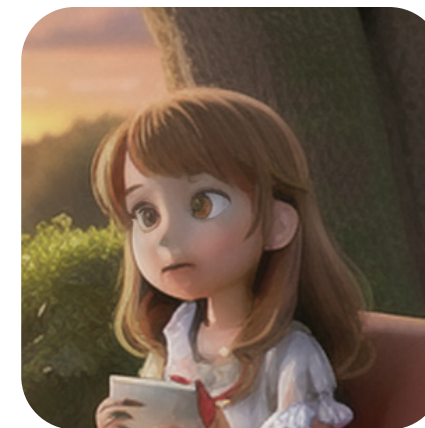
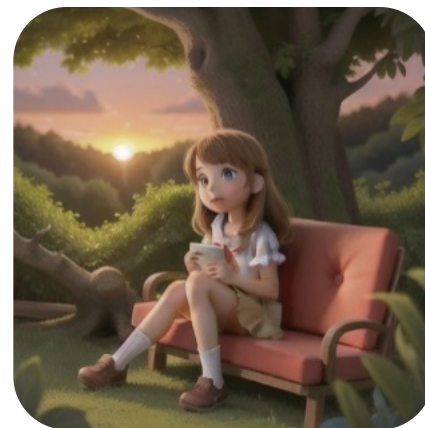
This is one of the simplest and very common use-cases using inpainting, to change minor aspects of a human's face or fix hands etc. Take note of the settings and steps I used to change her eye color to blue.





In Painting

Examples



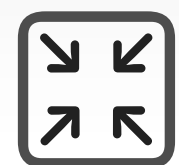
Changing eye color and adjusting symmetry



Adding a soccer ball to this cartoon 3D render of Woody

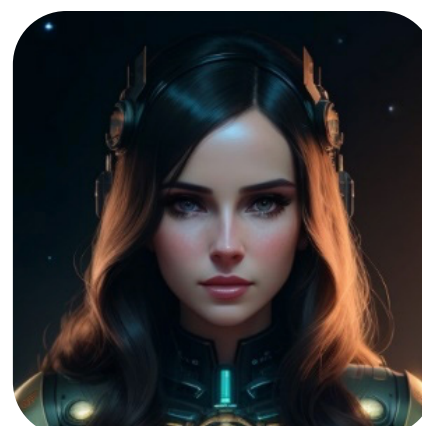
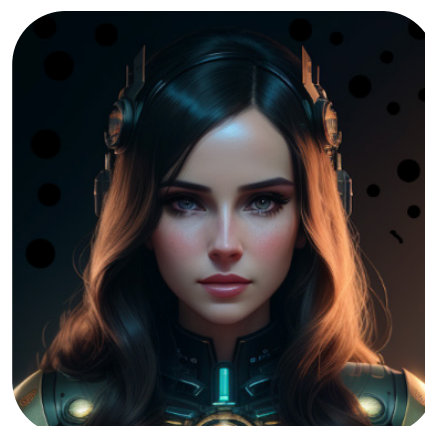
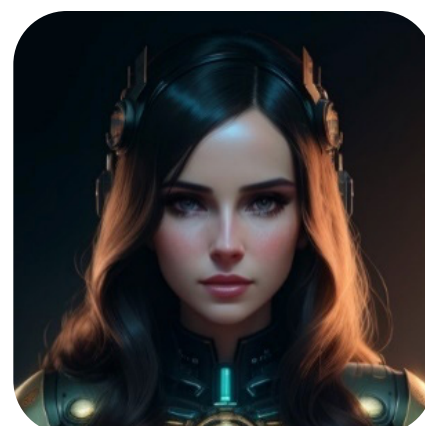


Changing the color of his boots after adding the soccer ball

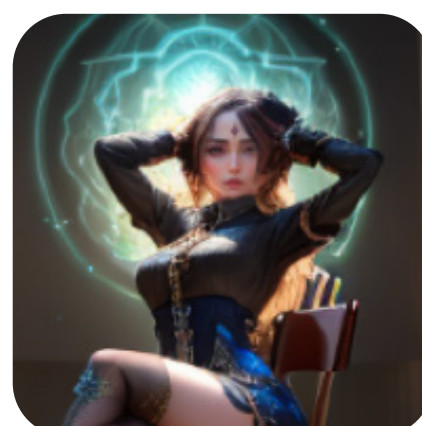
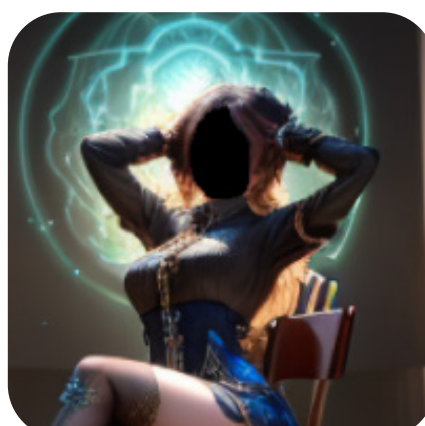
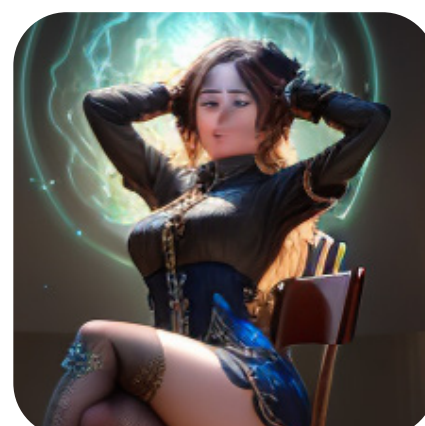


In Painting

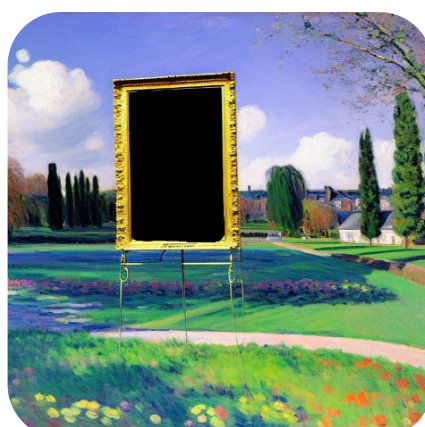
Examples



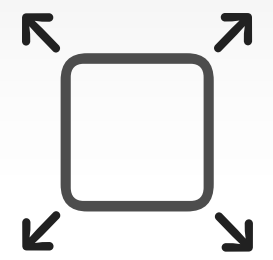
Adding tiny details like stars in the background



Fixing badly rendered facial feature



Changing entire parts of images



Out Painting



Outpainting is an exciting technique that leverages a large language model (LLM) to create new content that extends your existing image.

Belonging to the diffusion model family, it starts with a bit of random noise and gradually builds up detail to match your target image.

Stable diffusion, a specific type of diffusion model, is particularly good at generating high-quality images. It's trained on a rich dataset of images and text, which helps it master the art of relating images to their descriptions.

So, what can outpainting do for you?

- Expand Images: Extend your image in any direction - think panoramic views or additional details such as animals, objects, trees etc.
- Complete Images: Fill in the missing parts of an image.

But, remember, outpainting, like any other tool, has its quirks.

Your output depends on your input. So, a clearer, higher-quality input image will likely give you better results.

Being specific in your prompts also helps. The more detail you provide, the more likely the model will generate an image that fits your vision.

And don't forget patience is a virtue. Quality takes time, so give the model a few minutes to create your masterpiece.

Lastly, have some fun with it! Experiment with different settings and prompts to see the amazing results you can achieve with outpainting.



Out Painting

Let's dive into the key differences between Poor Man's Outpainting and Outpainting mk2

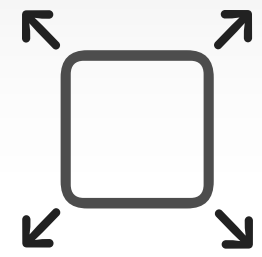
Poor Man's Outpainting: This is a more straightforward approach that uses a single diffusion model to generate the whole output image. It's less resource-intensive than Outpainting mk2, but it may yield lower-quality results.

Outpainting mk2: This technique employs two diffusion models to create the output image. The initial model sketches a rough outline of the output image, and the second one adds the details. Although this method might yield higher-quality results than Poor Man's Outpainting, it demands more computational resources.

Ultimately, your choice between these two techniques will depend on your specific requirements. If you prioritize speed and can compromise on image quality, Poor Man's Outpainting could be suitable for you. However, if you're aiming for high-quality results and are prepared for a longer wait, Outpainting mk2 might be your better bet.

Here's a summary of the key differences between Poor Man's Outpainting and Outpainting mk2:

Feature	Poor Man's Outpainting	Outpainting mk2
Diffusion model	Single diffusion model	Two diffusion models
Computational complexity	Less computationally expensive	More computationally expensive
Output quality	Lower-quality results	Higher-quality results



Out Painting

Outpainting mk2

To give the Outpainting mk2 method a spin, start by placing your image in the 'img2img' tab. Then, navigate to the 'Inpainting' subtab, click on the 'Script' menu, and select 'Outpainting mk2'.

As you get started, here are a few handy tips:

- Begin with just one or two outpainting directions. Try focusing on either just the left or right, or perhaps both left and right. Early on, it's advisable to avoid experimenting with top or bottom as it might yield unsatisfactory results.
- Once you're finished, make sure to revert the script back to 'none'. In many instances, you'll want to switch back to 'inpainting', and forgetting to close the Outpainting function might cause both features to run simultaneously.
- Begin with 'Pixels to expand' set at 256 and your 'Mask blur' option somewhere between 55-35.
- It's recommended to start with a low 'color variation' and a default of 1 for the 'Fall-off exponent'. As indicated in the UI, the lower this value, the more detail the model will attempt to include in your expanded image.

Masked content

☐ fill ☒ original ☐ latent noise ☐ latent nothing

Inpaint area

☐ Whole picture ☒ Only masked

Only masked padding, pixels 28

Sampling method

DPM++ SDE Karras

Sampling steps 20

☐ Restore faces ☐ Tiling

None

img2img alternative test

Loopback

✓ Outpainting mk2

Poor man's outpainting

Prompt matrix

Prompts from file or textbox

SD upscale

X/Y/Z plot

controlnet m2m

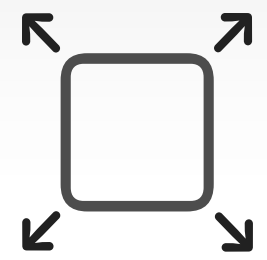
Outpainting mk2

Outpainting direction

☒ left ☒ right ☒ up ☒ down

Fall-off exponent (lower=higher detail) 1

Color variation 0.05



Out Painting

Outpainting Mk2

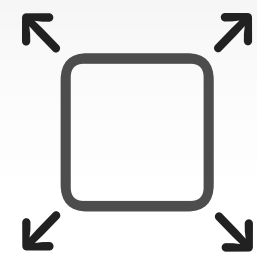
Next, I decided to expand this image. To do so, I experimented with the settings, as shown in the subsequent image. Please note that finding the right parameters often involves some trial and error.

With the adjusted settings, I obtained a satisfactory result: an expanded image with extensions on both sides.

Here's a pro tip: after using the outpainting function, consider reverting back to the inpainting function. This allows you to further refine the image according to your preferences. For instance, after just one quick attempt, I was able to transform the large tree trunk into a smaller, more proportionate tree.

To start with, I created this image using a straightforward prompt: "a photo of kids playing in the park". I employed the RealisticVision model and included some default negative prompts.





Out Painting

Poor Man's Outpainting

This method is relatively simple, let's head over straight to the img2img tab, click script: poor man's outpainting.
Tip: use the same seed as the original image.

Check out the settings in the image to to the right, these are my basic recommendations for an image with relative realism like this one.



Just resize **Crop and resize** Resize and fill Just resize (latent upscale)

Sampling method Euler Sampling steps 35

☐ Restore faces ☐ Tiling

Width 704 Height 704 Batch count 3 Batch size 1

CFG Scale 7.5

Denoising strength 0.65

Seed 2546967077

ControlNet v1.1.217

Regional Prompter

Script **Poor man's outpainting**

Pixels to expand 256

Mask blur 5

04

MASTER THE TOOL AND EXPLORE OUTWARDS

Alternative Checkpoints



In addition to the default Stable Diffusion models, the community has been actively engaged in creating customized and modified models. These models are trained on specific image databases, allowing for the creation of context-specific content that is often easier to generate compared to the official Stable Diffusion models (V1.4 through V2.1). Some of these custom models have gained popularity among users.

However, it's important to be aware of certain limitations when working with customized models. These models may have specific characteristics or biases based on their training data, which can affect the generated output. Bonus Tip: When downloading models, it is generally considered safer to choose models with “.safetensors” extensions over models with “.ckpt” extensions. This helps ensure a more secure process.

Examples of Various Models, Embeddings and Checkpoints

Notice how some of these examples have nearly identical prompts but change according to the model or embedding used.



Jeff Bezos, a mysterious magician, soft lighting, money in his hands

Via: RealisticVision Checkpoint



Caricature, happy white baby parrot, dressed as a gladiator, leather boots, standing in a very detailed colosseum, very detailed, intricate, anthro, cinematic light

Via: Anthro T.I Embedding



Tyrion Lannister, drinking wine.

Via: DisneyPixarCartoon Checkpoint



Caricature in pixar style, woody from toy story, dressed as a gladiator, leather boots, standing in a very detailed colosseum, very detailed, intricate, cinematic light

Via: DisneyPixarCartoon Checkpoint

How to Install:

First, start exploring different models on Image model and stable diffusion communities such as [Civitai](#), [Hugging Face](#) and [Discord communities](#).

Once you've found the model you'd like to install, click on the "download" button you find on the page such as on the screenshot from Civitai you see here.

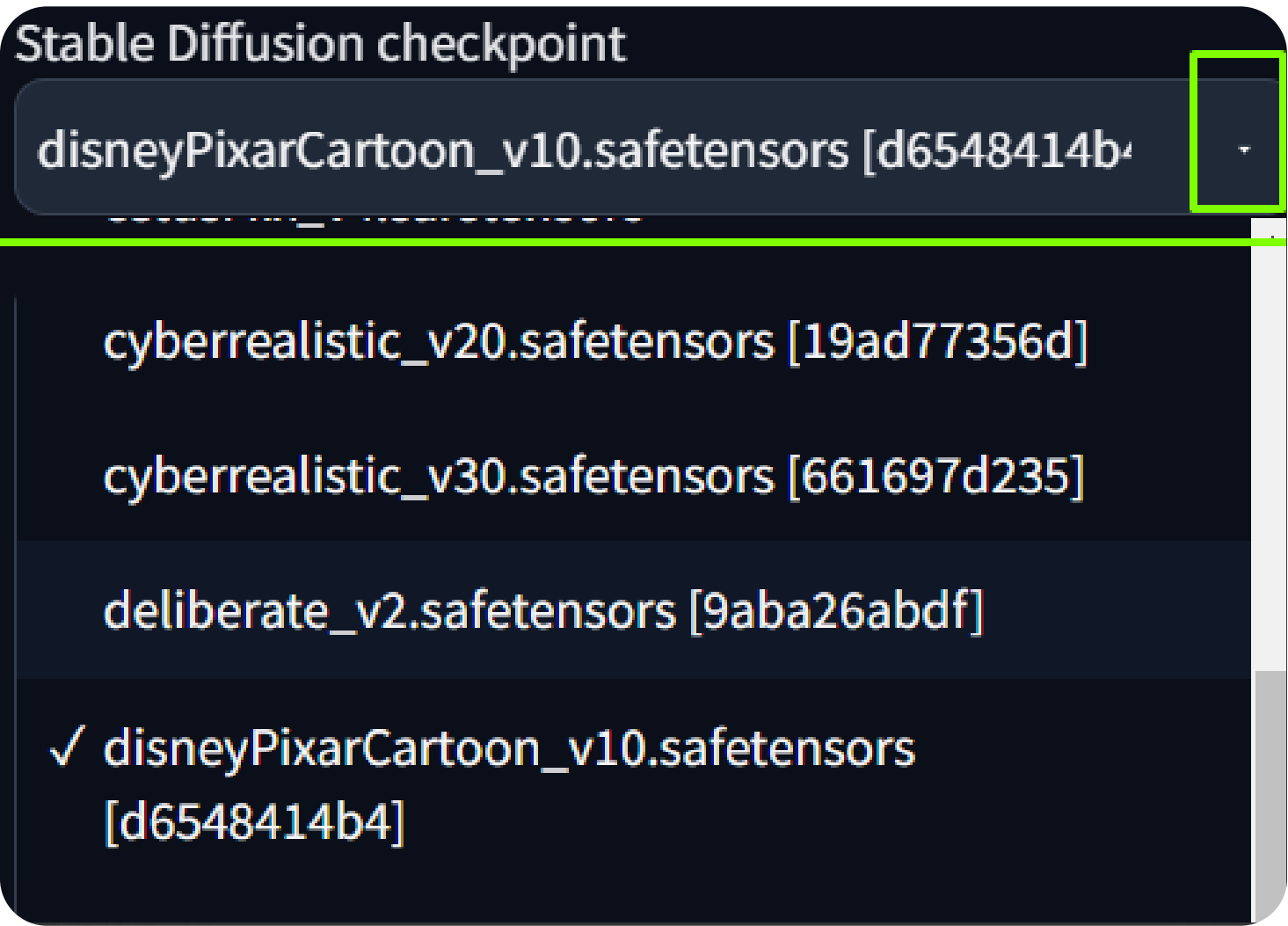
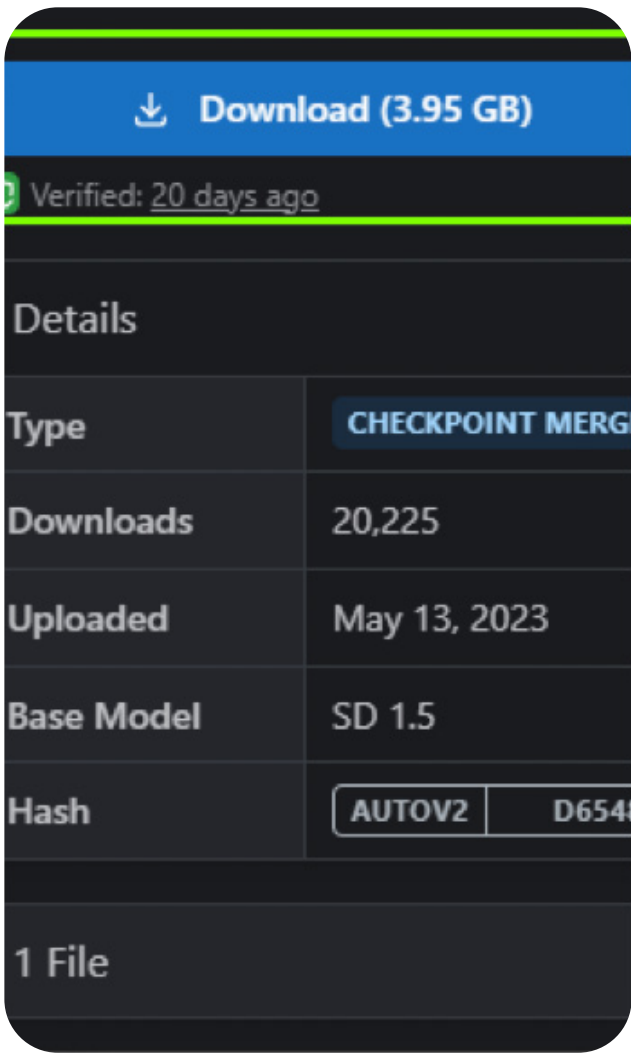
Once your file is downloaded, you only need to move it to the Stable Diffusion "models" folder.

You can do this by visiting the folder where you installed your webui in the installation phase. Then going to models\stable-diffusion

The address should look something like this:

C:\Users\{your username}\stable-diffusion-webui\models\Stable-diffusion

Here, you will paste your new model, then reboot your GUI if you had it open. Then you can click on the dropdown menu on the top left of the GUI and chose the model you want to render images ob.



Strengths and Limitations:

As with every extension, tool, modifier, embedding, or addition to your Stable Diffusion (SD) GUI, there are always advantages and disadvantages. Please refer to the following table for some fundamental pros and cons to keep in mind.

Pros	Cons
Capable of generating high-quality images spanning a broad range of styles and subjects.	The process can be time-consuming and computationally demanding, particularly when training custom models.
Useful in crafting realistic and captivating images for various applications.	There's a chance of generating images that appear blurry, pixelated, or unrealistic.
Provides a platform to explore diverse styles of art and design.	The models aren't flawless and may occasionally produce images that deviate from user expectations.



Examples

There is a large spectrum of different custom models out there.

Some of my personal favorite are: *Cyberrealistic, Deliberate V2, Realistic Vision, Dream Shaper, Disney, Open Journey, NED, GhostMix and F222.*

But it is important to note that you can find endless new models being created by the AI and Stable Diffusion community on Civitai, Hugging Face and Discord communities. I've had the pleasure of trying over 50 different models.

The biggest perk of using these models, is that even a beginner can easily write a very basic prompt and get great results.

In this example, I wrote what I call a "lazy prompt":

A photo of a group of creative artists, brainstorming in a modern office, hyperrealistic, super detailed.

Which is what a user would normally write on Midjourney. Although all versions had errors in the details and facial features, you can see that different models yield different results. Needless to say, not all models were intended for this use-case, but the example is to illustrate the differences.



Deliberate V2.0



Cyberrealistic V2.0



Stable Diffusion V2.1



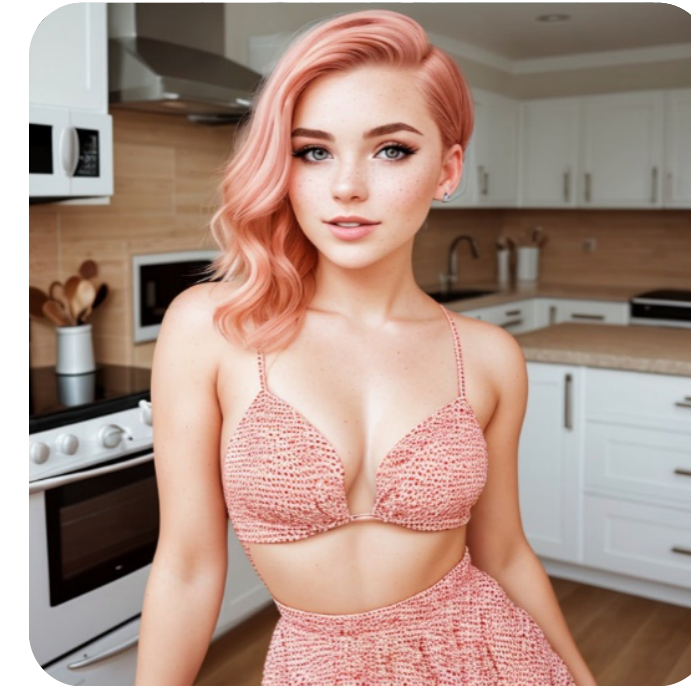
Realistic Vision V2.0

Examples: Deliberate V2.0

Now let's take a look what you can achieve when using the models for their intended styles.

First let's investigate one of my favorite models: **Deliberate V2.0**

I've found both by experimentation and by community recommendations that the best sampling method to use for this model is ***DPM++ 2M Karras***, but you can also use and experiment with ***DDIM*** and ***Euler a***.

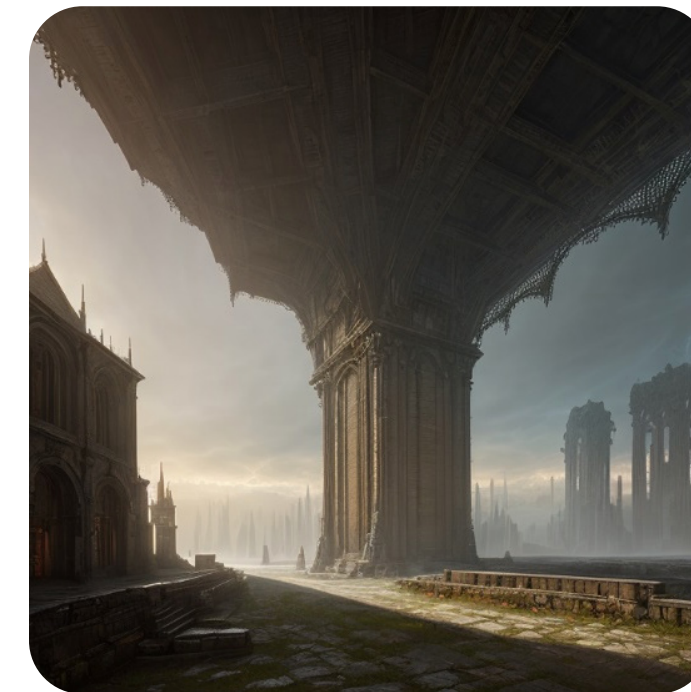
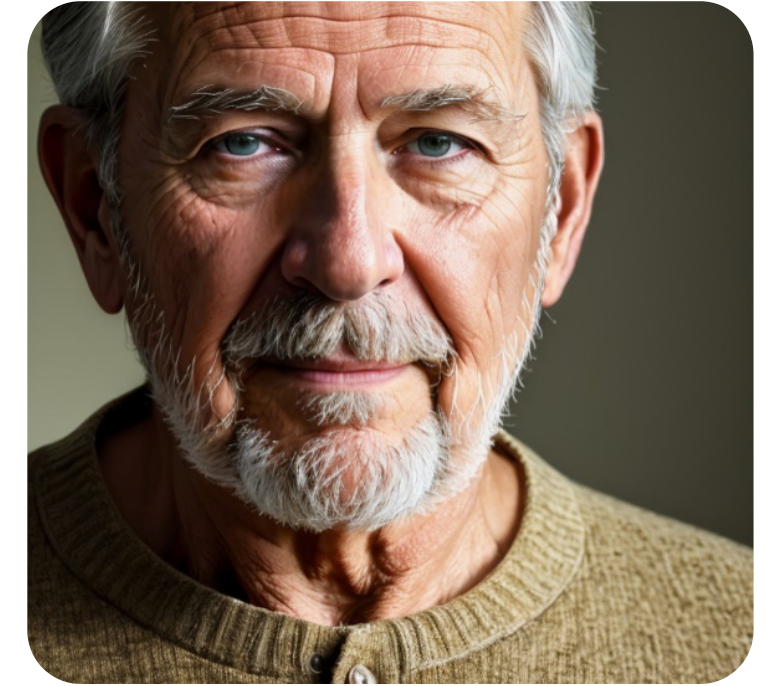
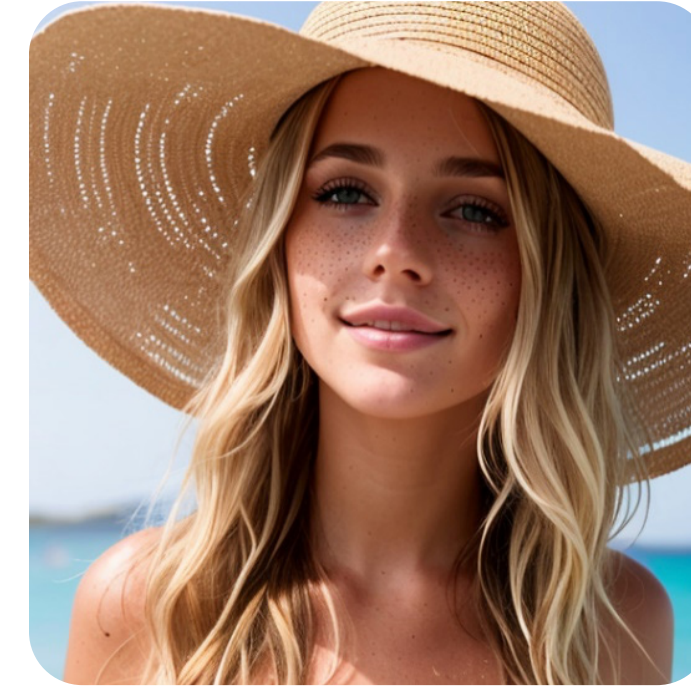


Examples: Cyberrealistic V3.0

Next let's take a look at one of the most fascinating realism-trained models: *Cyberrealistic V3.0*.

Recommended sampling method: DPM++ 2M Karras, DDIM and Euler a.

Recommended sampling steps: 20-35



Examples: Disney Pixar Cartoon

Next let me show you one of the interesting cartoon-based models: *Disney Pixar Cartoon Type A*

Note: Later on, I'll be showing you how you can achieve very cool similar results using an easier method called "Textual inversion embeddings"

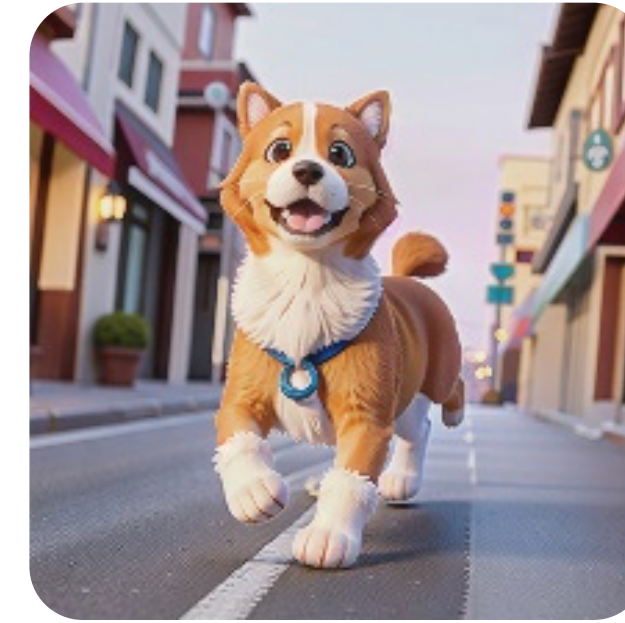
Recommended sampling method: Euler A.

Recommended sampling steps: 25-20

N.B: Some of these models are also trained on NSFW content, this is like the deep-web of Stable Diffusion, so be cautious in your prompts. An easy pro tip to avoid this is to always add this into your negative prompt: *(nudity, nude, naked, NSFW:1.5)*



More Disney Examples



Potential Drawbacks

(LORAs and finetuned models):

Limited Flexibility: Utilizing pre-trained or finetuned models can be limiting as they may confine your work to a particular style or approach. They are not specifically designed to understand or interpret your unique needs or requirements.

Limited Recognition: Many models may fail to identify or process certain keywords effectively if they were not included during the training phase. For example, if you present an image of a flower vase with the description "Photo of (Lion:1.2) on a couch, flower in vase, depth of field, film grain, Fujifilm XT3, crystal clear, 8K Ultra HD, dark studio", some models may not be able to recognize and properly react to all of the components of this complex prompt.

Legal and Ethical Risks: Using pre-existing models can increase the risk of infringing upon copyrights or inadvertently crossing ethical boundaries in terms of content usage. It is crucial to fully understand the terms and limitations set by the providers of these models to prevent any potential legal or ethical violations.

These drawbacks are relatively less noticeable with textual inversion embeddings since they're essentially just a bunch of lines of text embedded into your prompt.



LORA

LORAs, an acronym for Low-Rank Adaptation of Large Language Models, aren't traditionally viewed as extensions of stable diffusion models. They function by adjusting the weights in various models.

While often employed together with stable diffusion models, their use isn't limited to them.

In essence, LORAs tailor stable diffusion models to align better with certain themes or concepts, such as art styles or characters. They give us the flexibility to modify the model's behavior to meet specific demands.

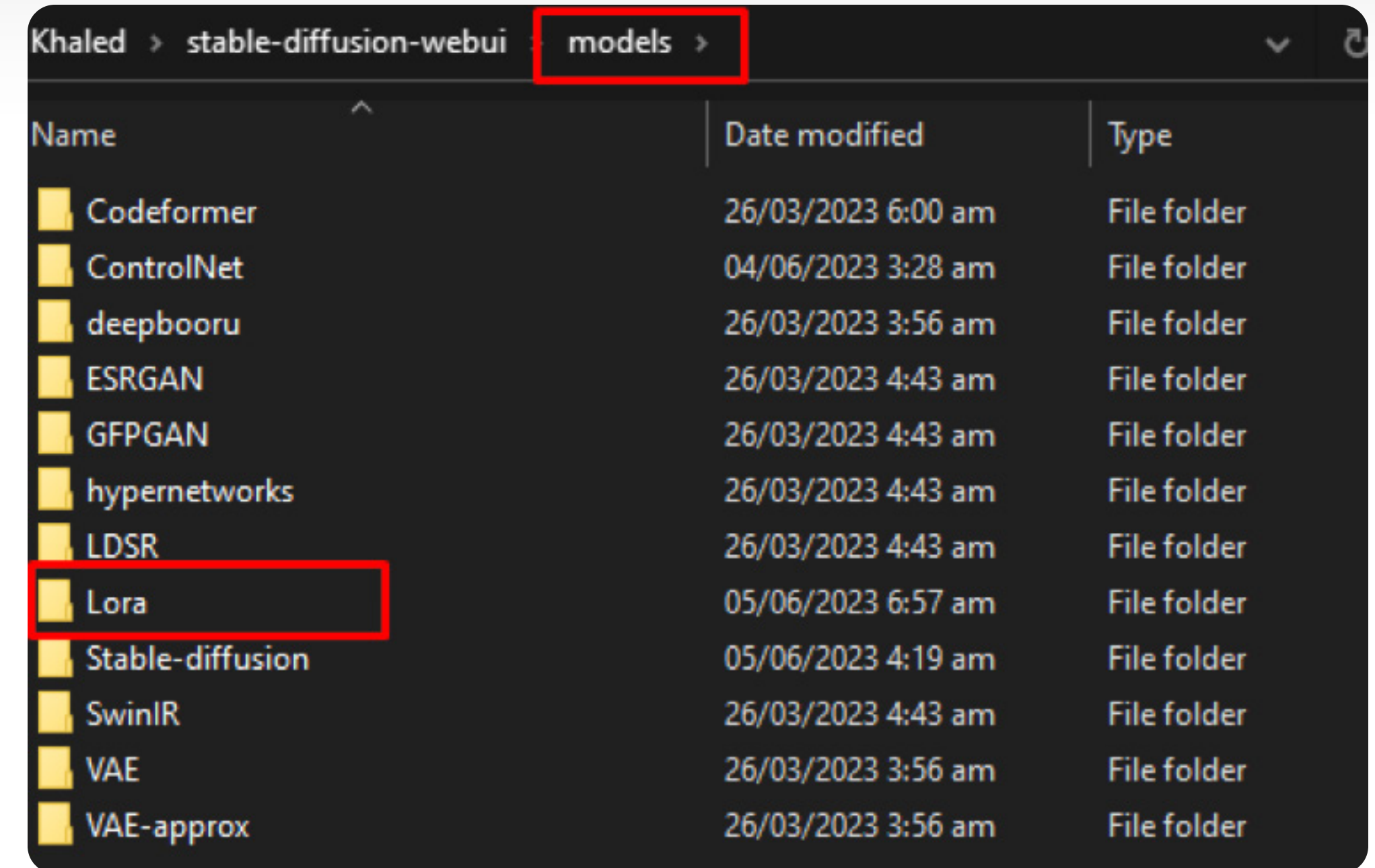
In simpler terms, LORAs, coupled with straightforward prompts, can produce stunning results. Additionally, with thoughtful prompting, you can shape entire concepts around a LORA that are theme-specific.

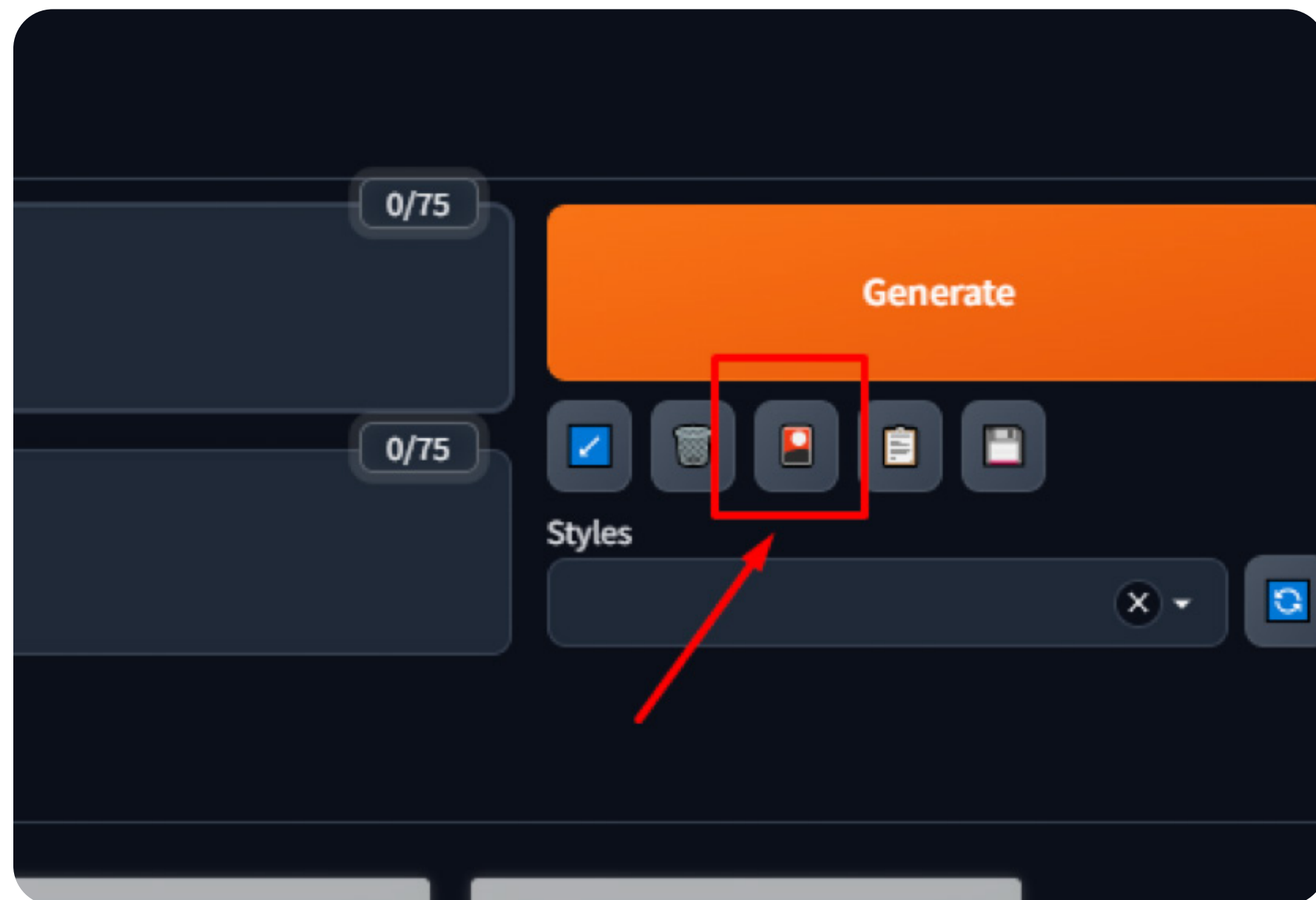
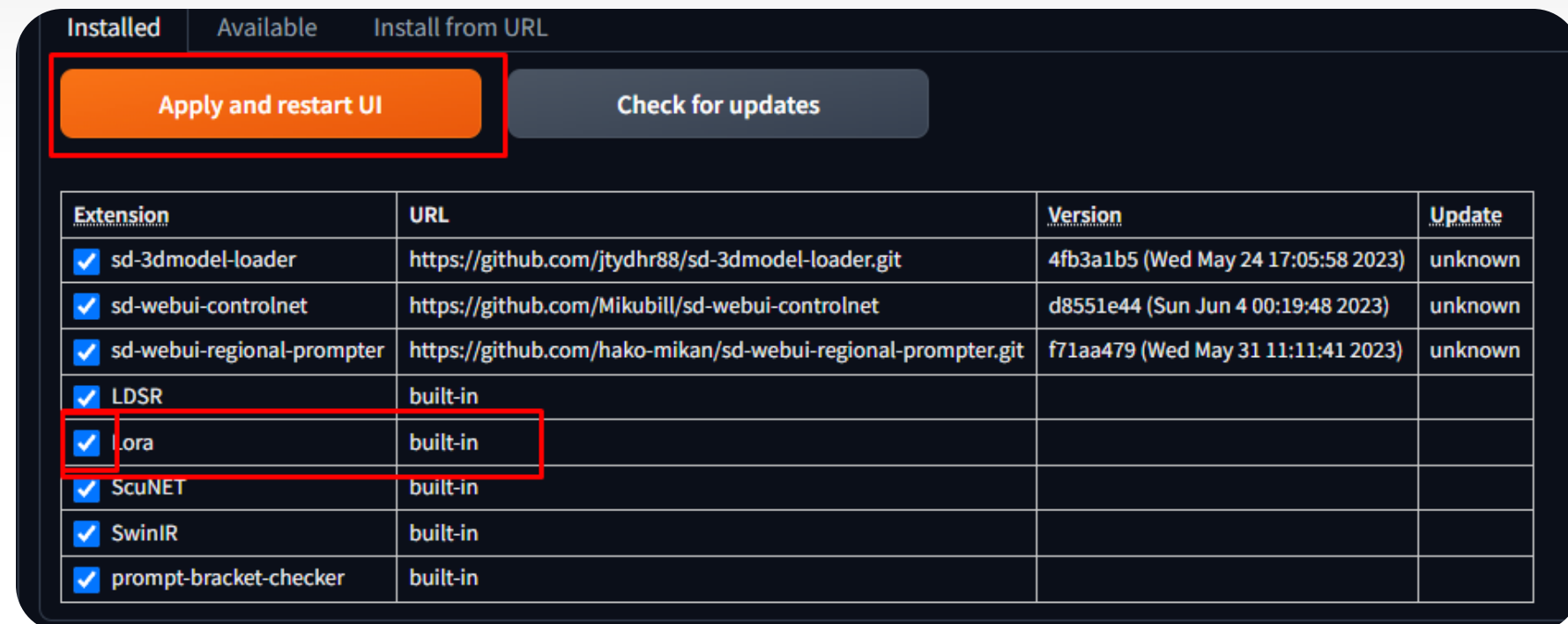


How to Install

The easiest way is to visit a stable Diffusion community website such as Civitai.com, filter your search with “LORA” and find one you like, download the file, copy and paste it into your LORA folder, which is usually found in your webui “models” folder.

C:\Users%USERNAME%\stable-diffusion-webui\models\LORA





How to Use a LORA

First make sure you have the LORA extension enabled in your GUI, by heading over to the extensions tab, checking the LORA checkbox then apply and restart the UI

Next, you need to head over to your GUI, click on the little sunset icon beneath the “generate” button to see your LORA library.

In simple theory, to use a LORA, you need to add <LORA file name:numerical factor from 0 to 1>

For example: <LORA:steampunkai:1>

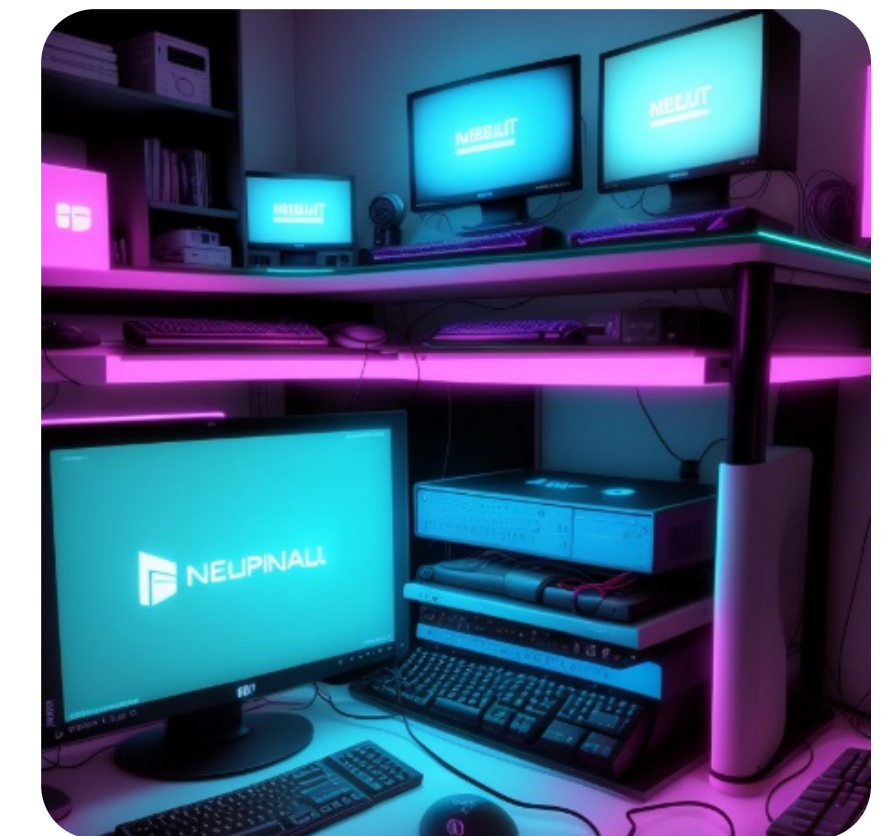
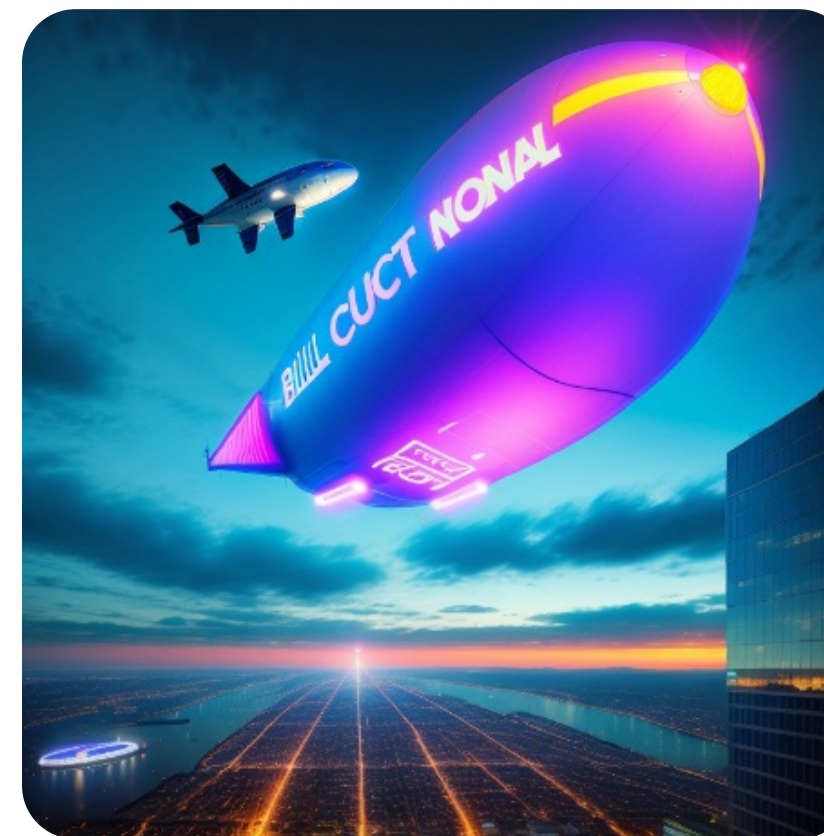
LORA Examples

Neonpunk LORA

By adding the following Lora trigger in my 3 original prompts:

`<lora:Neonpunkai-8-000002:0.6>`

I got these results without changing any other parameters.



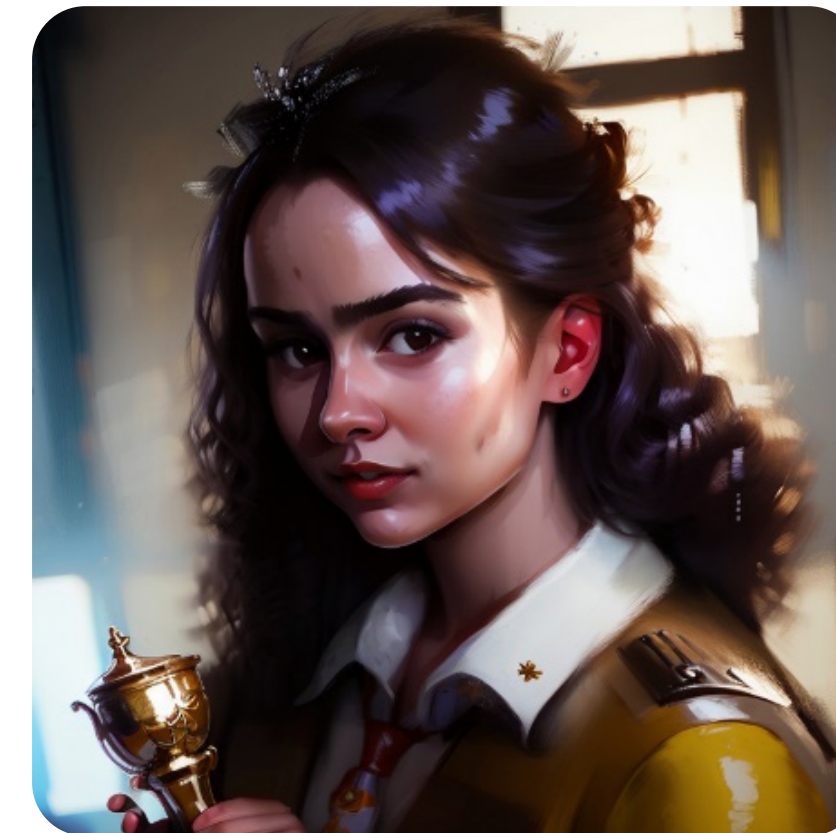
LORA Examples

There are tones of examples, from LORAs that help get perfect hands in images generated, to ones that associate with color, clothing, poses, specific real people, landscape, modelling, architecture and a lot more.

These are a few “before and after” examples for different LORAs



Medieval Armor LORA



Artistic Style LORA



3D LORA

Stable Diffusion Extensions

Stable diffusion extensions encompass supplementary tools or methodologies created to augment the efficiency and functionality of established diffusion models.

In this book, we will explore one of the most noteworthy extensions, ControlNet. It's fair to assert that ControlNet represents the most valuable tool marketers can leverage within Stable Diffusion.





ControlNet

ControlNet is an extension for Stable Diffusion that enables users to control the composition and human poses in generated images. It achieves this by creating an image map from an existing image, which serves as a guide for generating new images.

It can also help you with a lot of other things not just human poses, in a more general sense, it enables you to use a reference image to create more accurate end results in your prompts.

You can either create/source these reference images yourself or use preprocessing extensions that will help generate the reference images for you.

Examples of input references you can provide ControlNet to control your output include Canny edge, M-LSD lines, Human Pose, Scribbles, Depth map and Semantic segmentation.

Not to mention you can customize and create your own control variables, but we won't dive deep into that in this book.

ControlNet empowers users to create realistic and captivating images, making it a valuable tool for marketing and advertising purposes.

ControlNet Installation

Step one:

Open your cmd, navigate to the folder where you have your webui installed.

You can do this by typing in the following:
`cd {folder location}`

In my case:

`cd C:\Users\{User name}\stable-diffusion-webui`

Then type in the following command and wait till the file is installed.

`pip install opencv-python`

```
Command Prompt
C:\Users\Nabil Khaled>C:\Users\Nabil Khaled\stable-diffusion-webui
'C:\Users\Nabil' is not recognized as an internal or external command,
operable program or batch file.

C:\Users\Nabil Khaled>cd C:\Users\Nabil Khaled\stable-diffusion-webui
C:\Users\Nabil Khaled\stable-diffusion-webui>pip install opencv-python_

Command Prompt
C:\Users\Nabil Khaled>C:\Users\Nabil Khaled\stable-diffusion-webui
'C:\Users\Nabil' is not recognized as an internal or external comm
operable program or batch file.

C:\Users\Nabil Khaled>cd C:\Users\Nabil Khaled\stable-diffusion-we
C:\Users\Nabil Khaled\stable-diffusion-webui>pip install opencv-py
Collecting opencv-python
  Using cached opencv_python-4.7.0.72-cp37-abi3-win_amd64.whl (38.
Requirement already satisfied: numpy>=1.21.2 in c:\users\nabil kha
ackages (from opencv-python) (1.24.2)
Installing collected packages: opencv-python
Successfully installed opencv-python-4.7.0.72

[notice] A new release of pip available: 22.2.1 -> 23.1.2
[notice] To update, run: python.exe -m pip install --upgrade pip

C:\Users\Nabil Khaled\stable-diffusion-webui>
```


ControlNet Installation

Step TWO:

Start your webui

go to the extensions tab, click on the install from URL tab

Paste the following URL (or search for it on Google by typing “latest github repository for controlnet automatic1111”

<https://github.com/Mikubill/sd-webui-controlnet.git>

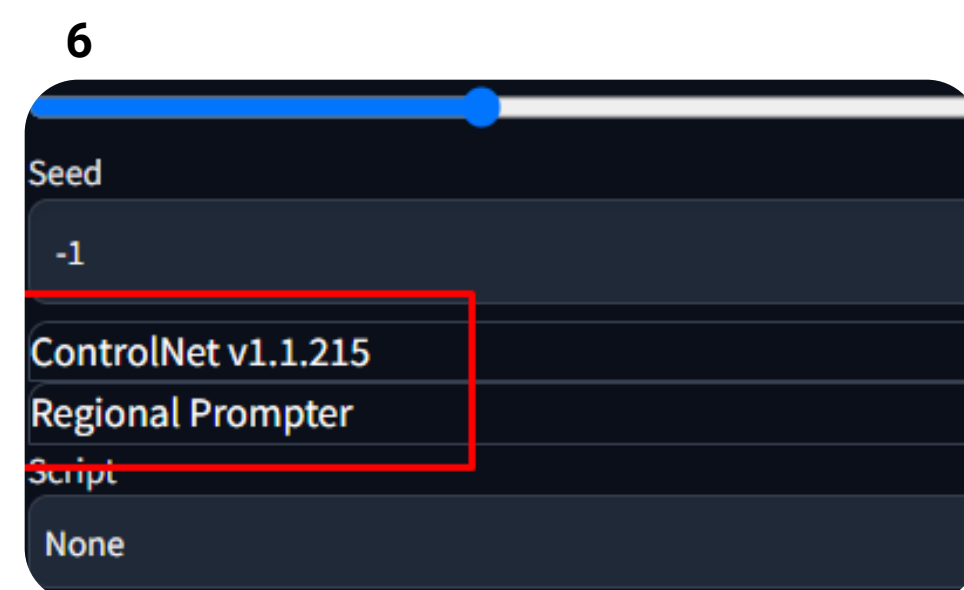
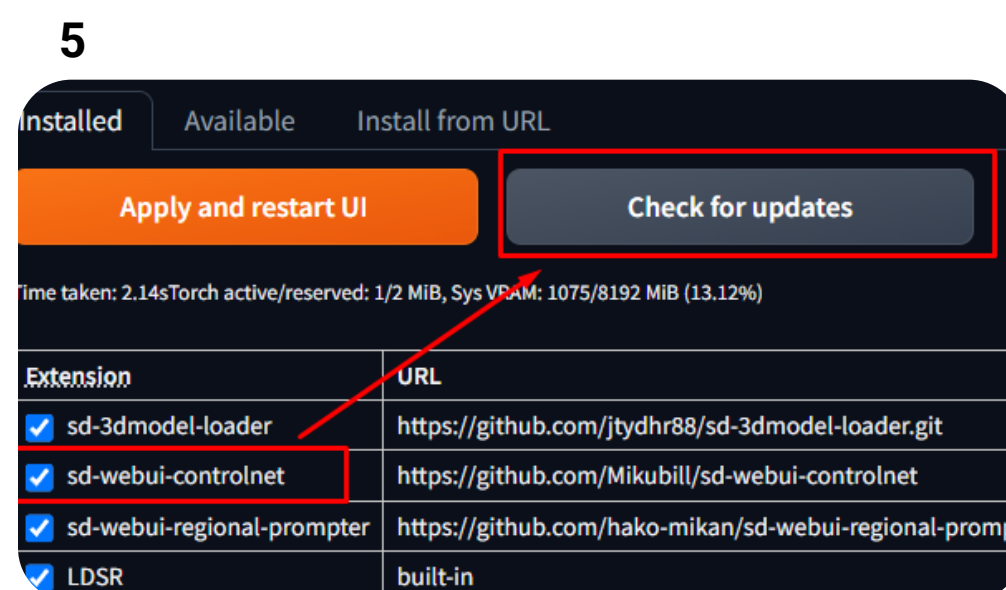
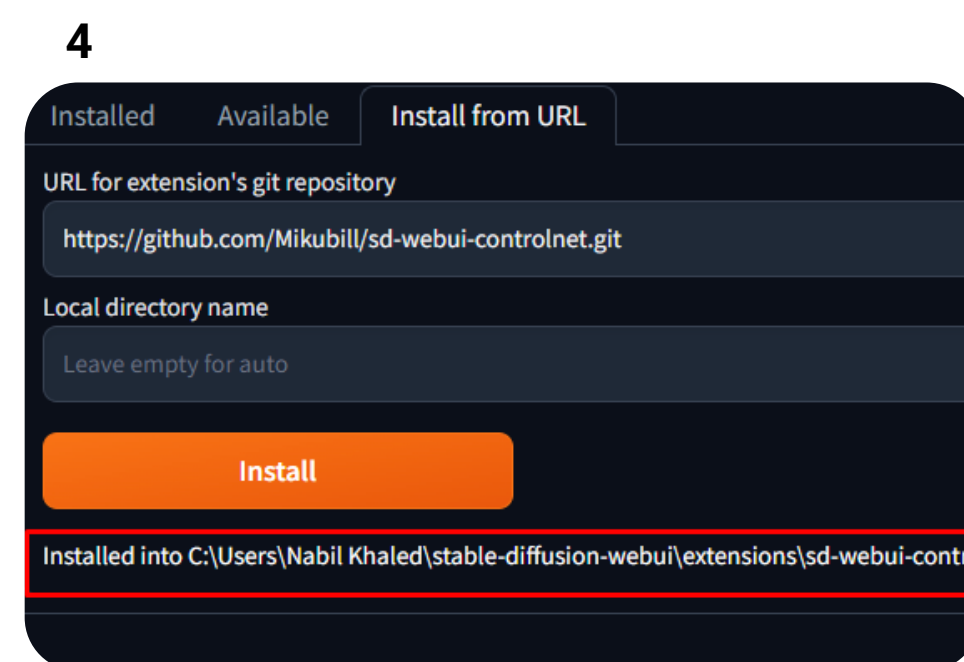
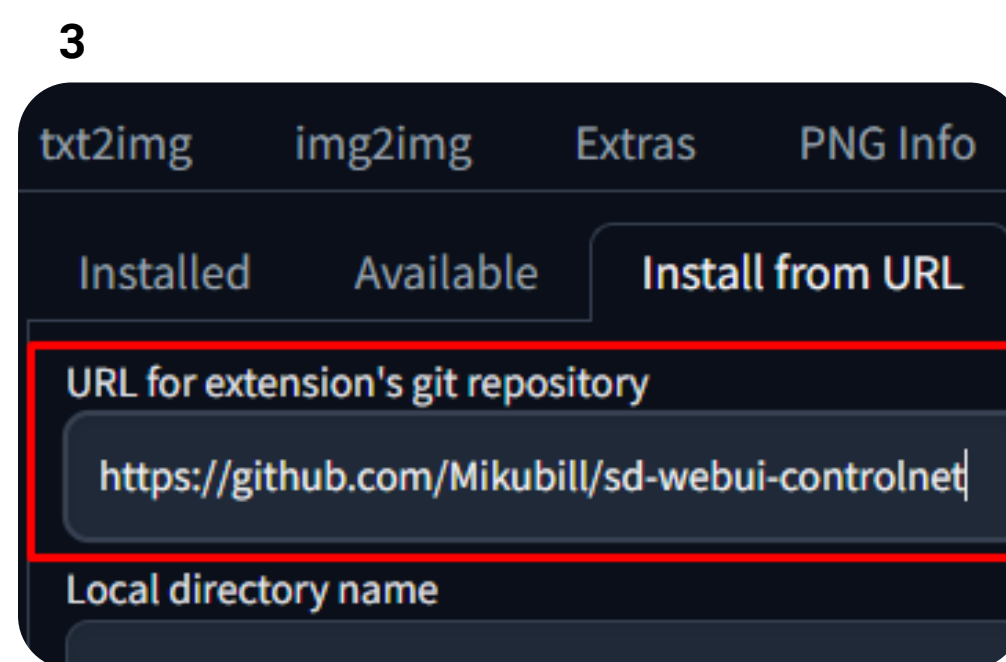
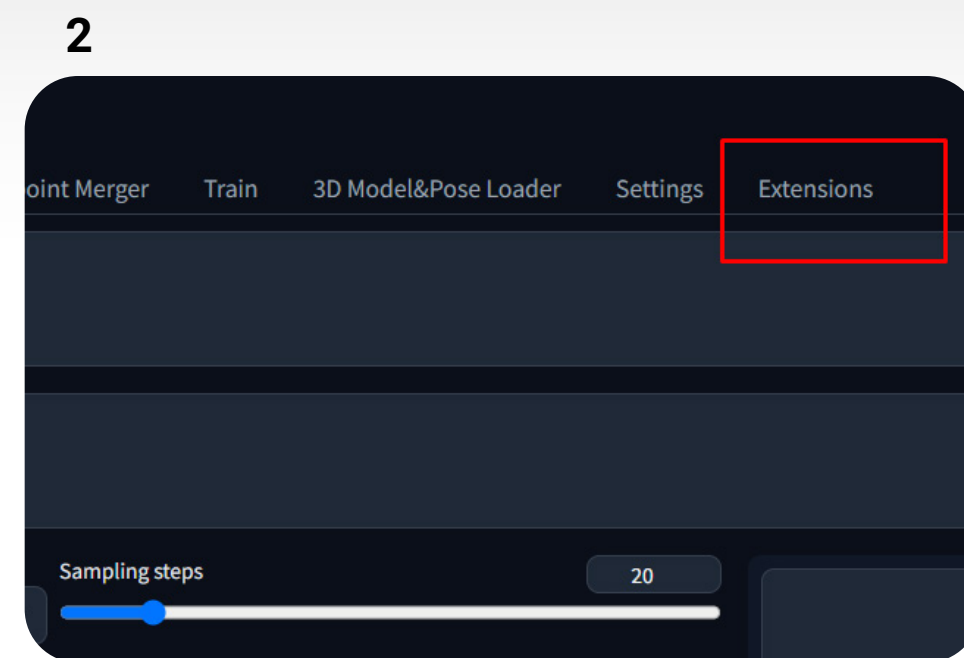
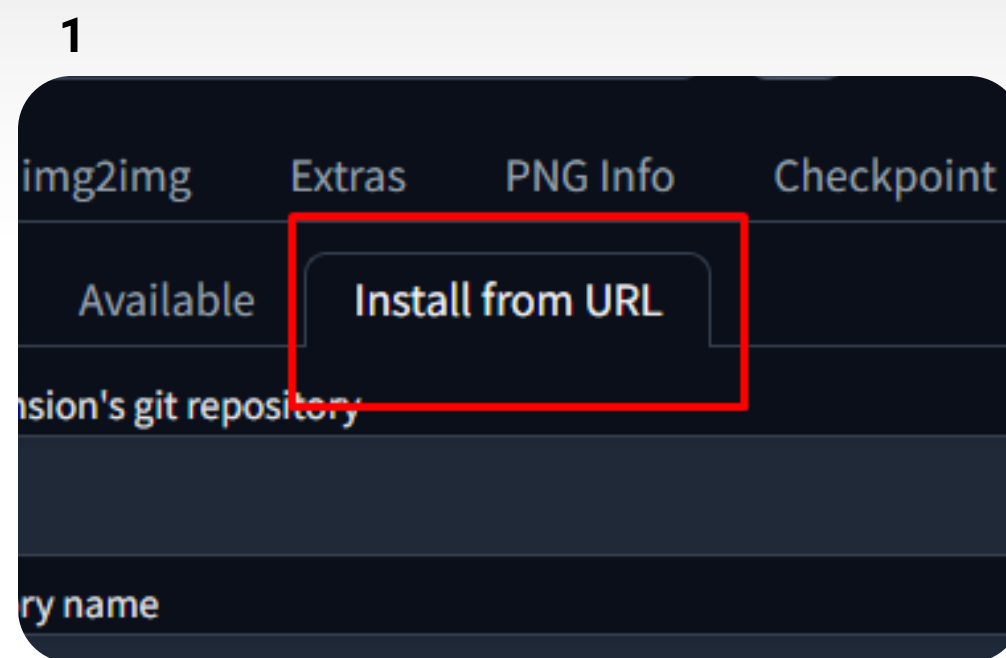
Once it is installed, head over to the “installed tab” and click on the checkbox for the newly installed controlnet extension

Then click “check for updates”

Then “apply and restart UI”

Note: You may or may not need to shut down your webui entirely and reopen it manually.

You know it's installed once you see “ControlNet” then the version right under your “Seed” tab



ControlNet Controls

As previously stated, there are various “control types” you can employ. Some customized ControlNet models even allow the simultaneous application of multiple control types.

Pay special attention to the “preprocessor” dropdown menu and feel free to explore and experiment with the various options available.

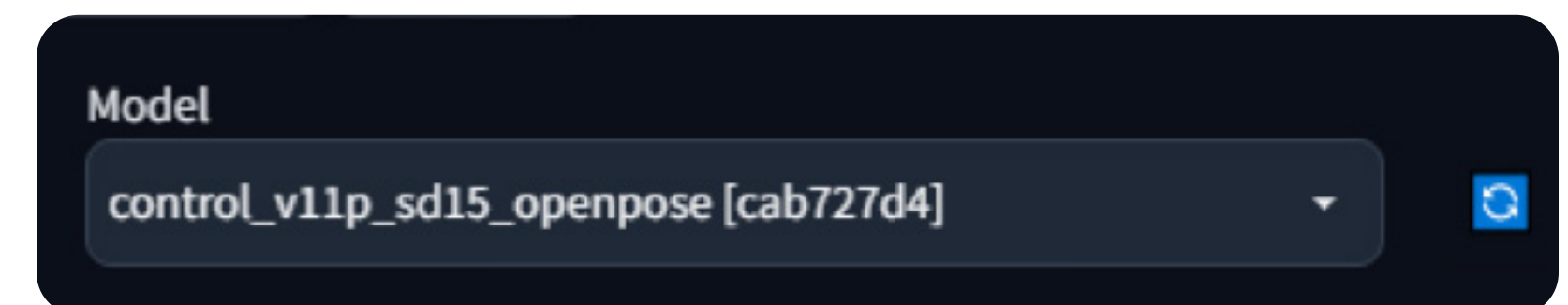
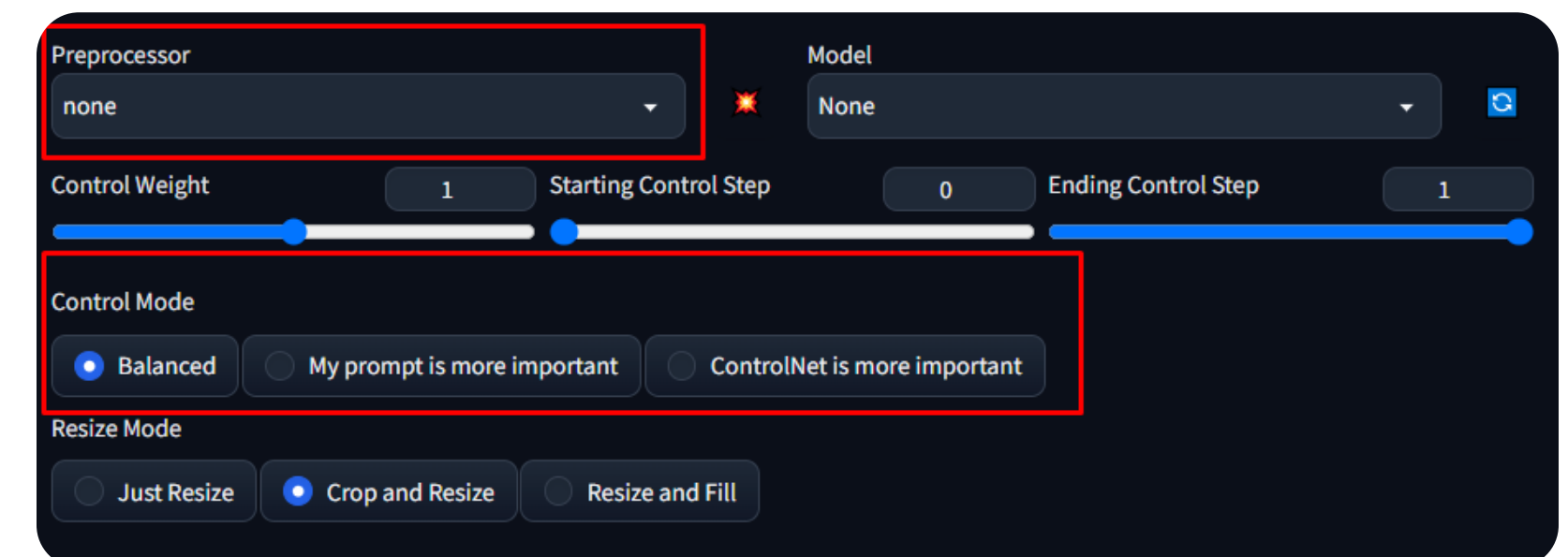
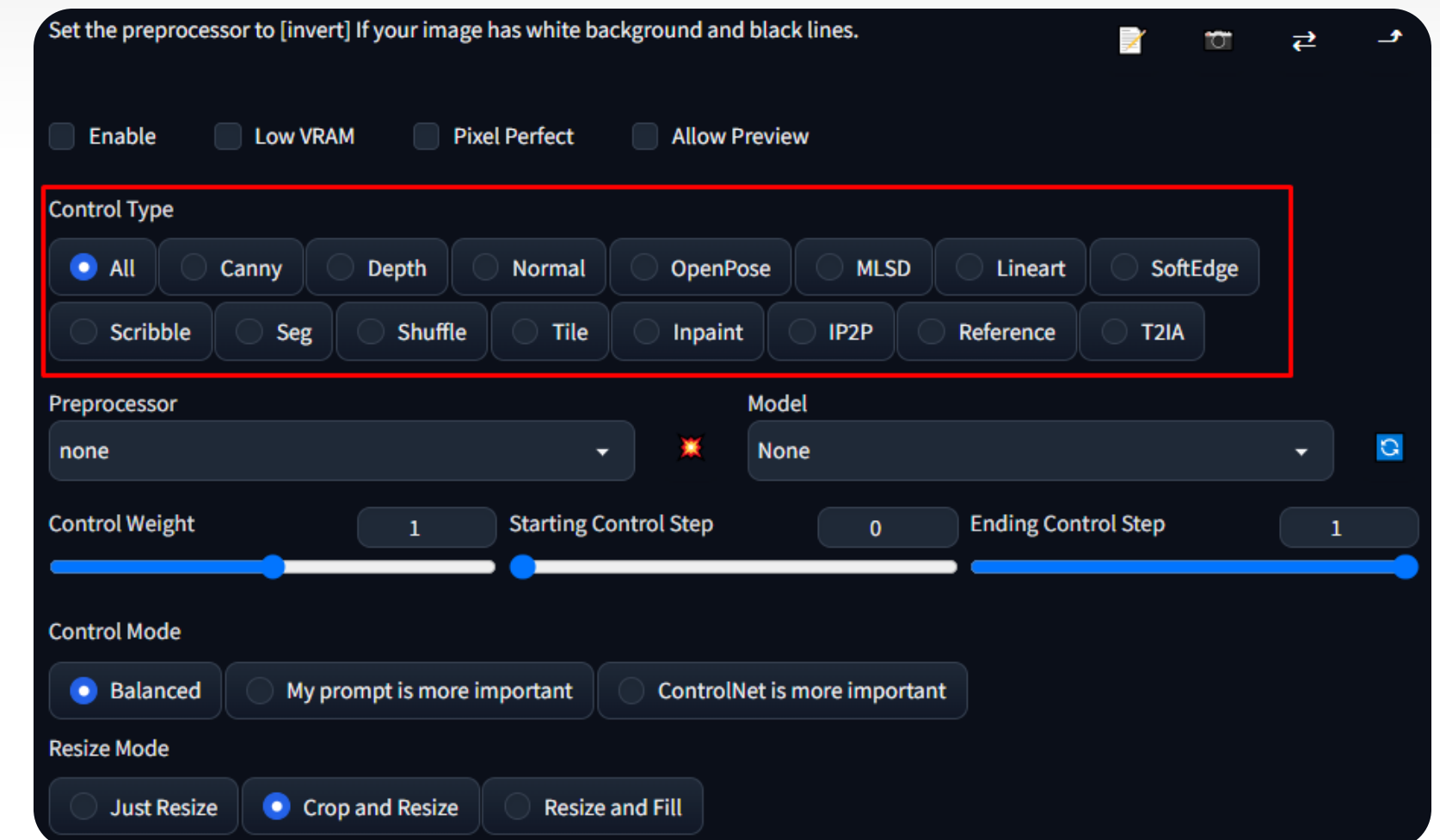
Another essential component is the “Control Mode.” This is straightforward: the option “my prompt is more important” gives priority to your prompt, “ControlNet is more important” prioritizes ControlNet input, and “Balanced” evenly weighs both.

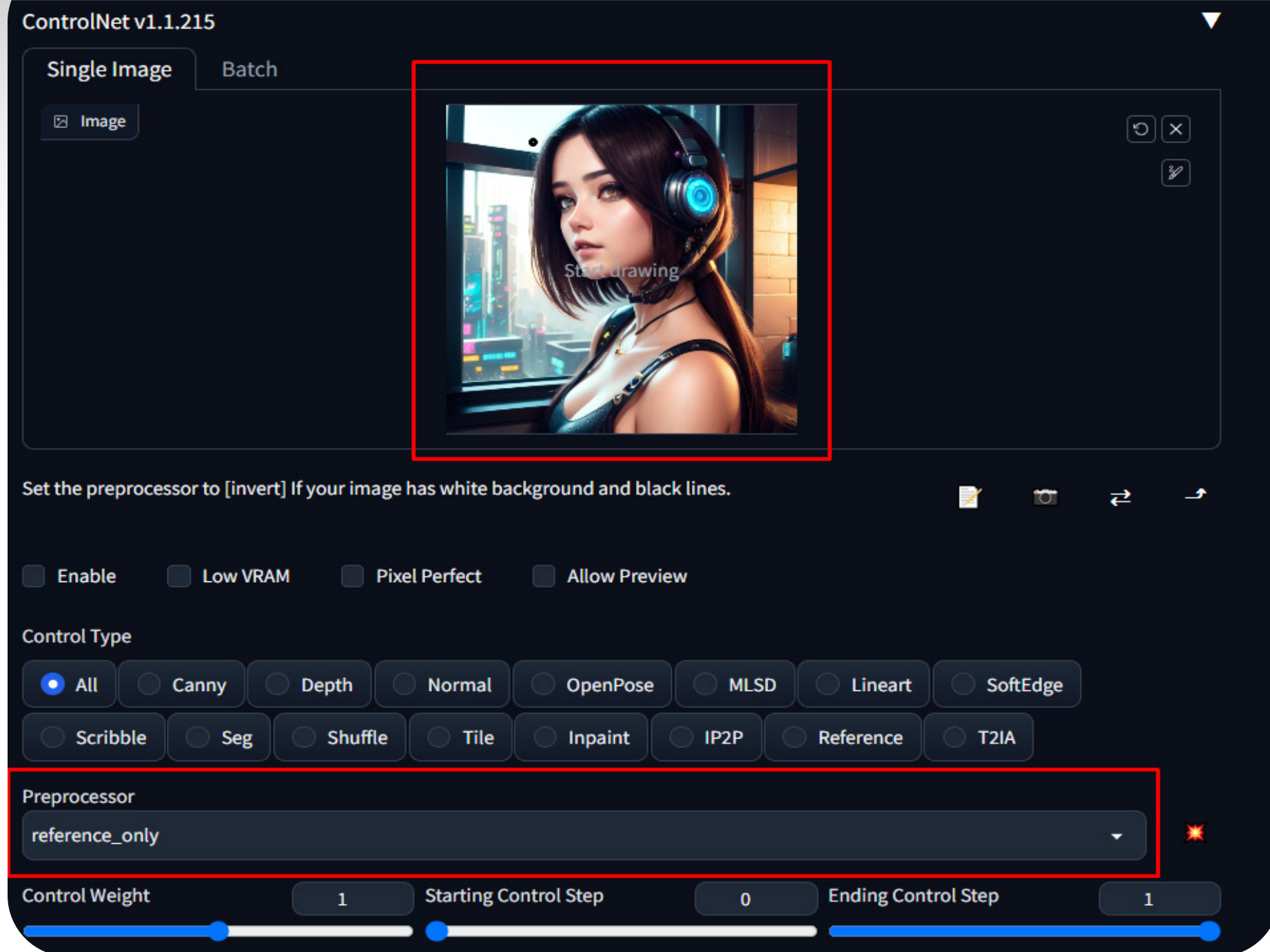
To achieve better outcomes, remember to download the ControlNet models, available here: [ControlNet Models](#). It’s advisable to understand the process before diving in; you can find a detailed guide here: [ControlNet Guide](#).

For a visual demonstration, you might find Sebastian Kamph’s video helpful: [Video Tutorial](#)

Once you’re done and you’ve installed the ControlNet models, go ahead and pick the model that ends with “openpose”

Keep in mind, this technology evolves at breakneck speed, and the information can become outdated quickly. Always seek the most recent updates from reliable sources like GitHub and Hugging Face.



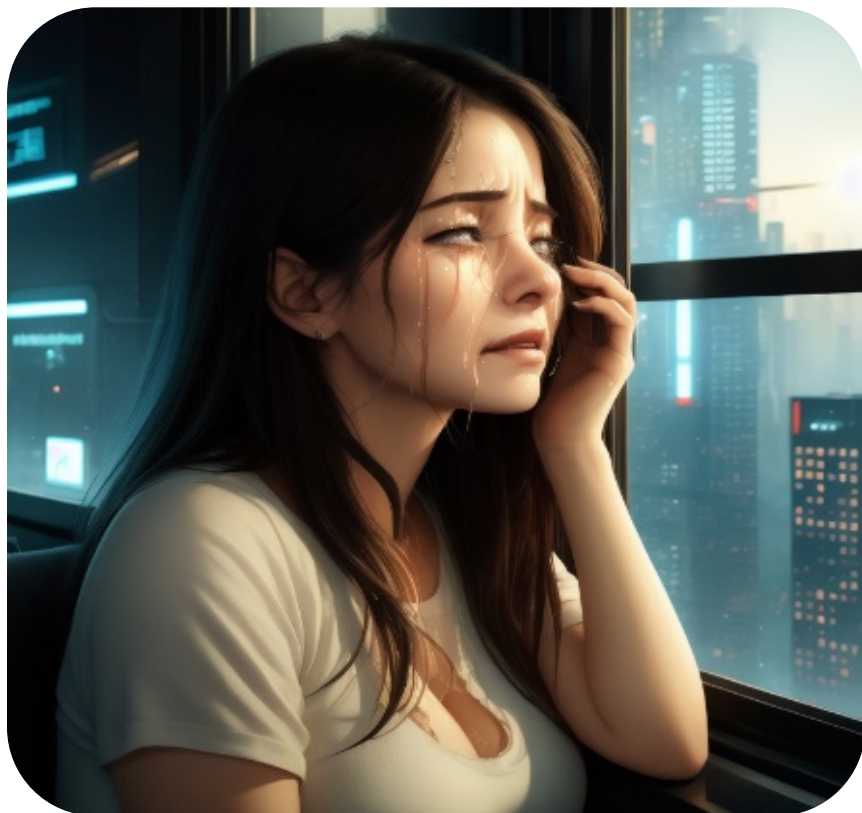


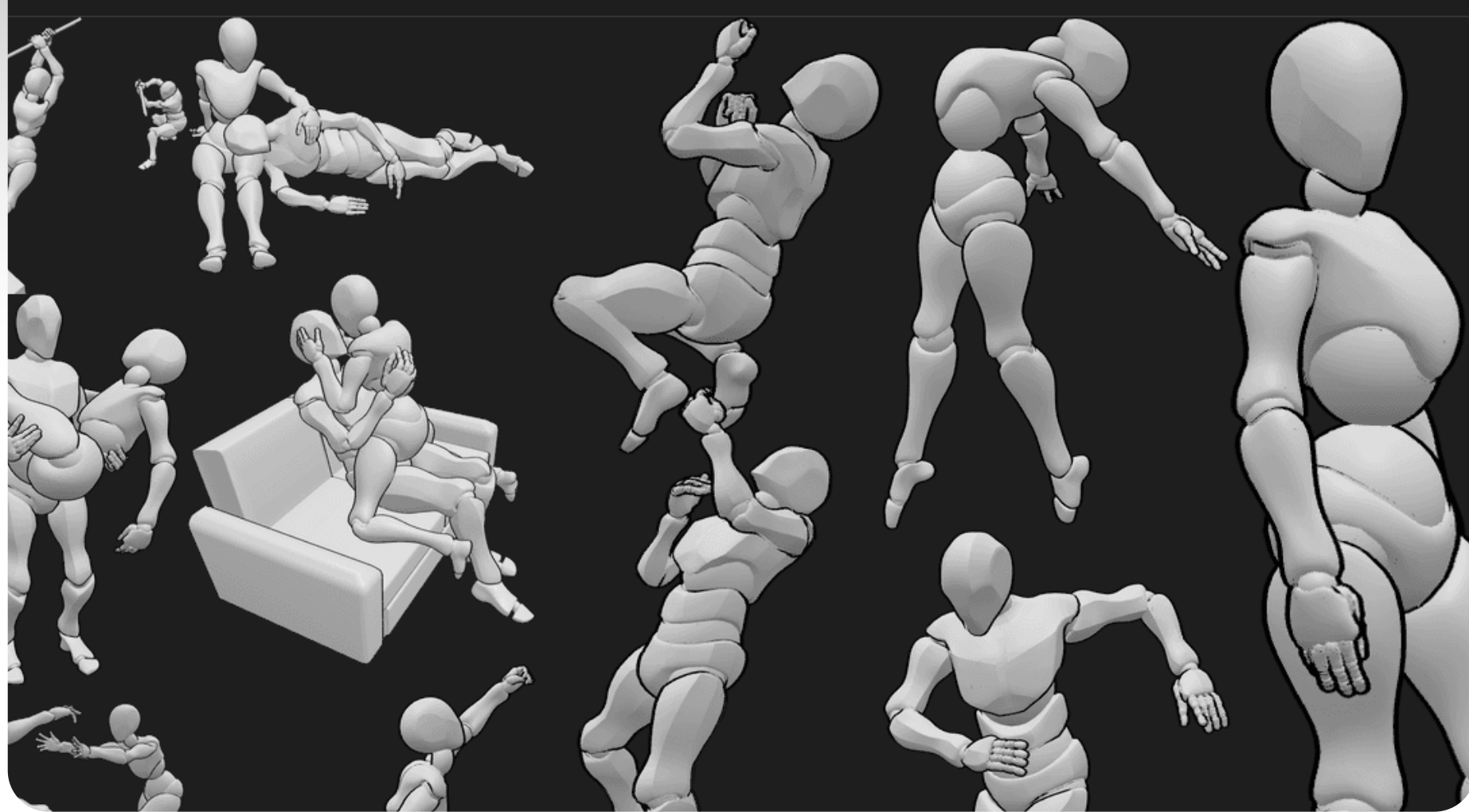
ControlNet Use-Cases

Let's delve into a recent, innovative preprocessing option known as 'reference_only'. I experimented with it using an image I had generated via Stable Diffusion. Following this, I incorporated a text prompt, 'woman crying:1.3'.

To make this work, I relied on the same Stable Diffusion model that I used to create the original image, in conjunction with previously saved negative prompt preset styles.

The outcome was quite good, but not as great as I would have hoped. After some tweaking, I was able to get a better result.





ControlNet Use-Cases

Human poses

First, we need to generate a pose.
You can go to <https://posemy.art/> or any other alternative.

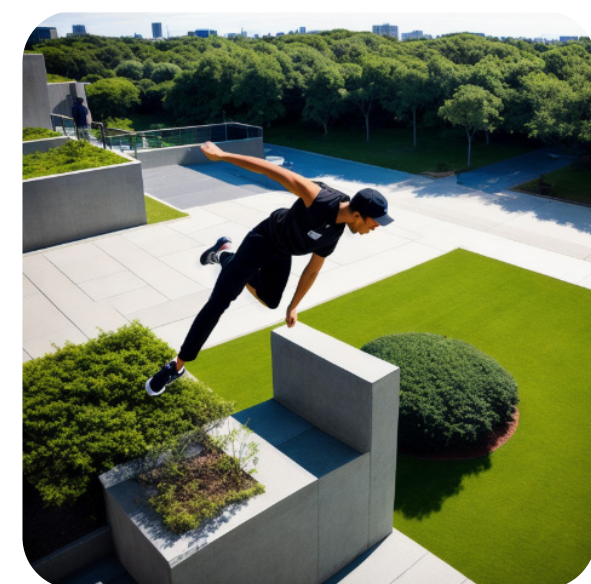
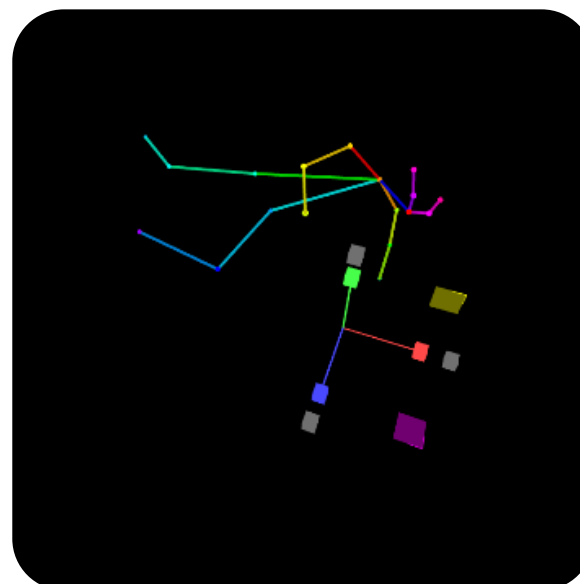
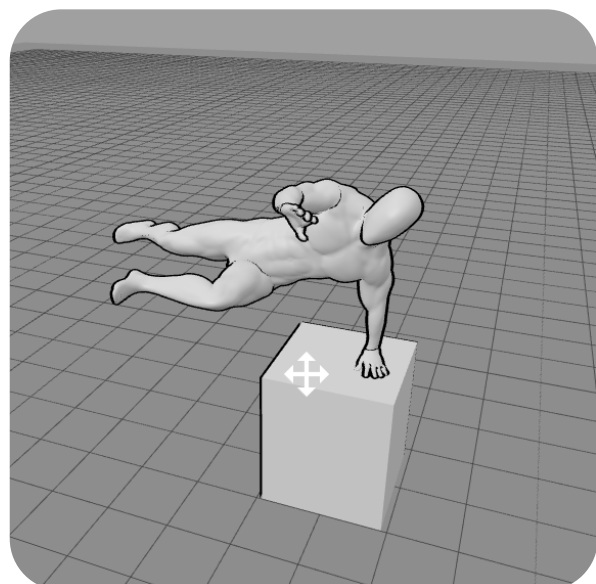
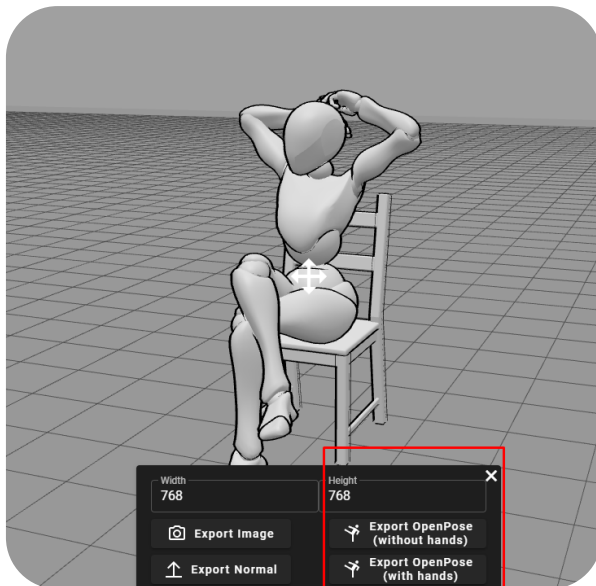
For the purpose of this example, I picked a “premade scene” you can easily find them on the top left of the window inside the pose my art UI.

Then export, and you will end up with a pose file with a black background and a pose “map” which is basically a bunch of color-coded lines.

You can then take that pose file and add it into ControlNet, set your control type to “Open Pose”

And then experiment and play around, with different poses, prompt styles etc.

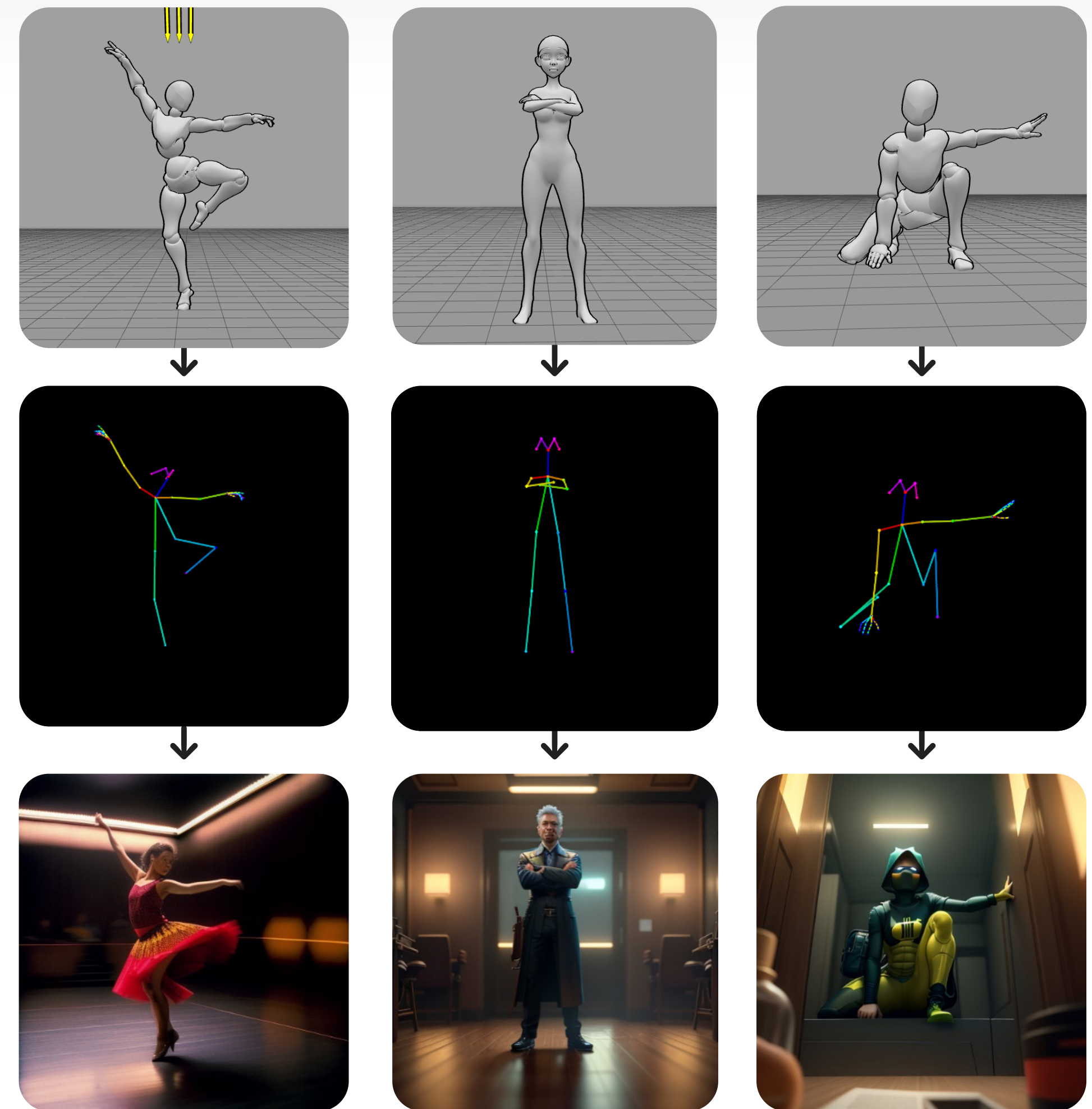
For the purpose of these examples, very simple prompts were used.



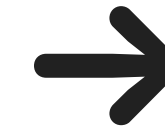
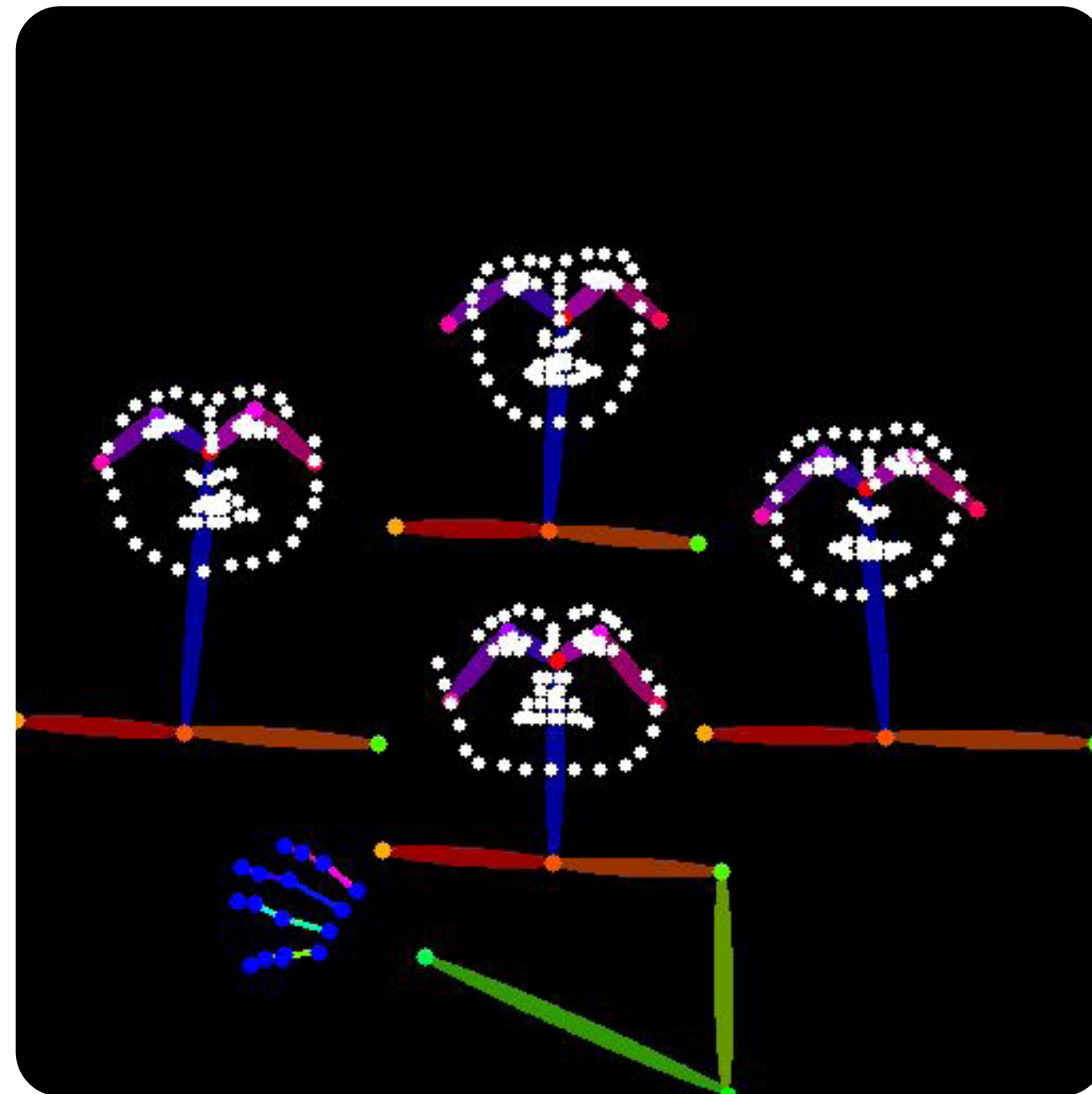
OpenPose:

Take a look at the various human poses you can achieve using <https://app.posemy.art/> and ControlNet

Some of the output images has some issues, but that was the prompt not the extension, as far as the pose and the control factor goes, it works seamlessly, although the ControlNet model is currently in Beta



OpenPose:



ControlNet Use-Cases

The following presents a prime illustration of how ControlNet can enhance your marketing collateral.

In this specific scenario, I employ the 'canny-edge' model.

By leveraging the outline image provided, in conjunction with the settings displayed in the Graphical User Interface (GUI) screenshot, we can expertly capture an impeccable image of the whiskey bottle. This includes all the meticulous details such as the label, brand, and detailed information imprinted on the bottle itself.

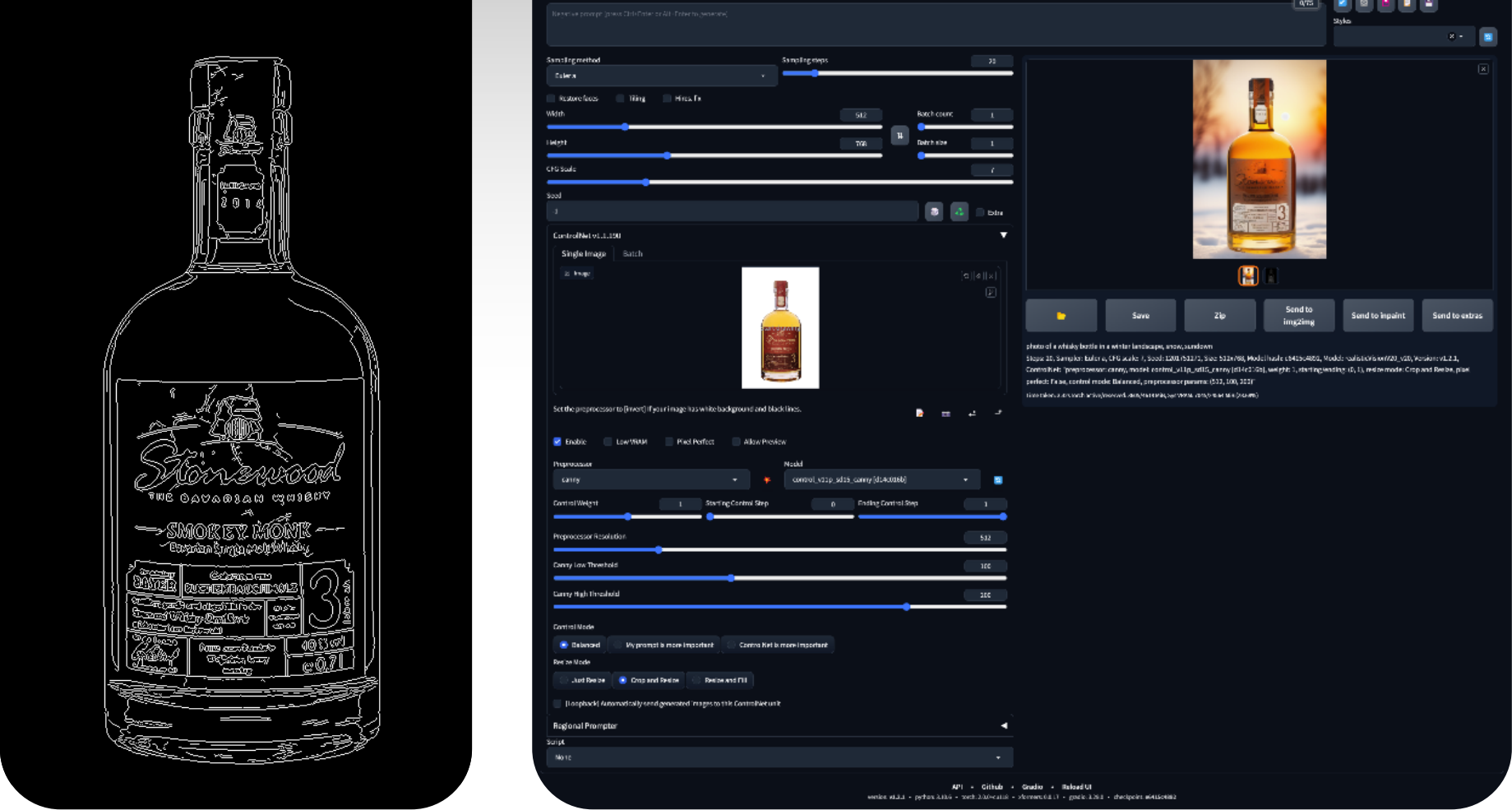


Photo of a whisky bottle in a winter landscape, snow, sundown
Steps: 20, Sampler: Euler a, CFG scale: 7, Seed: 1201751271, Size: 512x768, Model hash: e6415c4892, Model: realisticVisionV20_v20, Version: v1.2.1, ControlNet: "preprocessor: canny, model: control_v11p_sd15_canny [d14c016b], weight: 1, starting/ending: (1 ,0), resize mode: Crop and Resize, pixel perfect: False, control mode: Balanced, preprocessor params: (200 ,100 ,512)"

ControlNet Use-Cases

Logo Brainstorm

Using very simple sketches, which you can on paint, photoshop, illustrator or even procreate, you can give ControlNet the guidance you need to in order to visualize and conceptualize new logos, or graphic elements, or images of any sort.

In this case, I went ahead and tried creating a concept of a 3D snake logo, using a -30second sketch on Windows Paint, then again to create an “S” in the shape of a shark, which was also guided using a -30second Paint sketch of a shark.

☒ Enable

☐ Low VRAM

☐ Pixel Perfect

☒ Allow Preview

☐ Preview as Input

Control Type

☐ All

☒ Canny

☐ Depth

☐ Normal

☐ OpenPose

☐ MLSD

☐ Lineart

☐ SoftEdge

☐ Scribble

☐ Seg

☐ Shuffle

☐ Tile

☐ Inpaint

☐ IP2P

☐ Reference

☐ T2IA

Preprocessor

canny

Model

control_v11p_sd15_canny [d14c016b]

Control Weight

0.9

Starting Control Step

0

Ending Control Step

1

Preprocessor Resolution

512

Canny Low Threshold

100

Canny High Threshold

200

Control Mode

☐ Balanced

☐ My prompt is more important

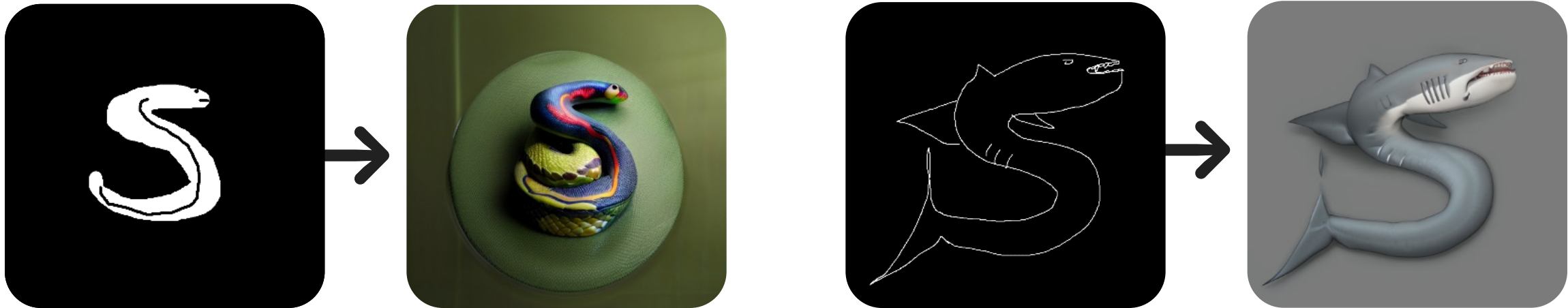
☒ ControlNet is more important

Resize Mode

☐ Just Resize

☒ Crop and Resize

☐ Resize and Fill



Try this Out

Extracting poses

First, go to your ControlNet tab under img2img, chose an image of a person in any pose you want to extract.

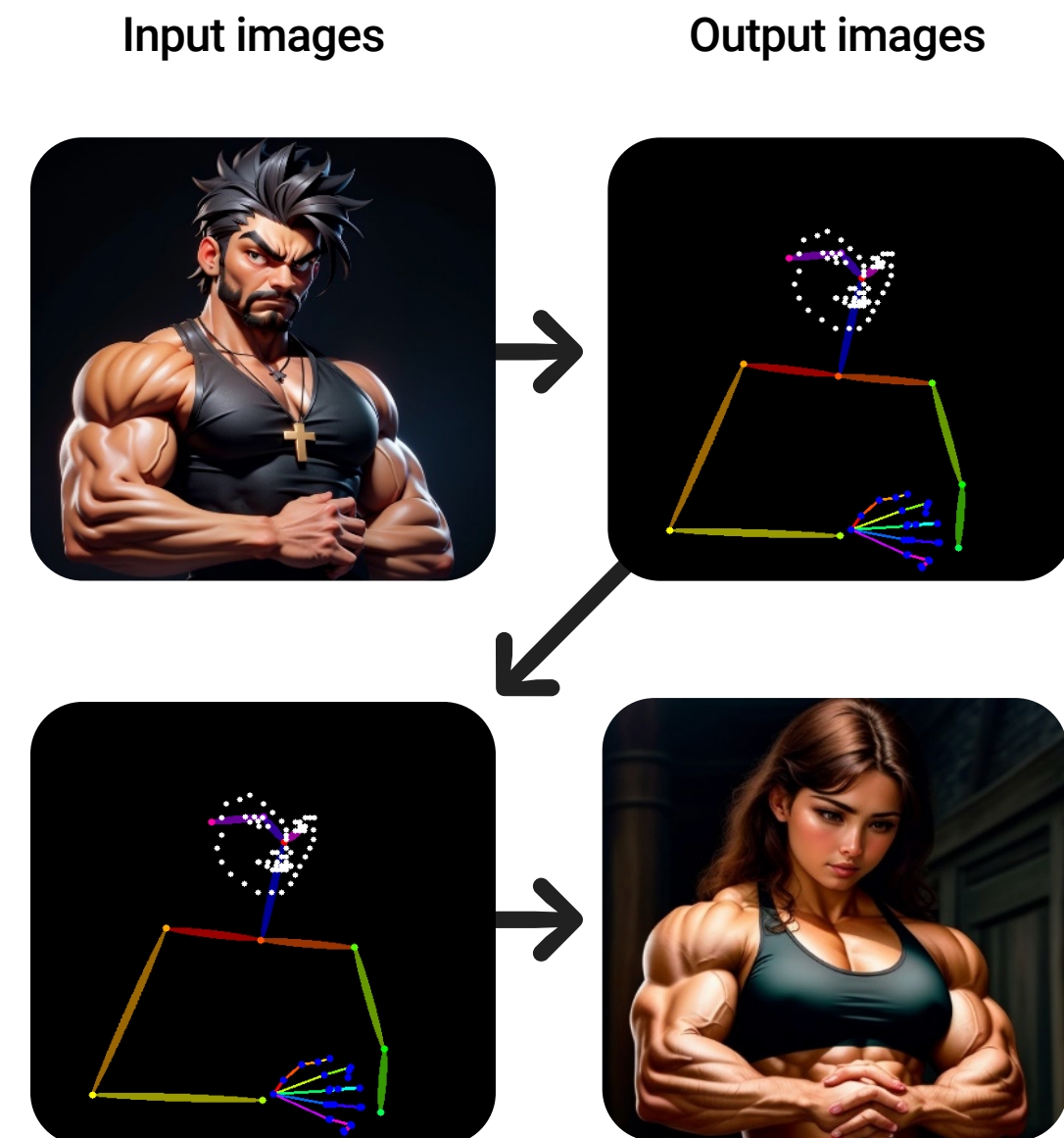
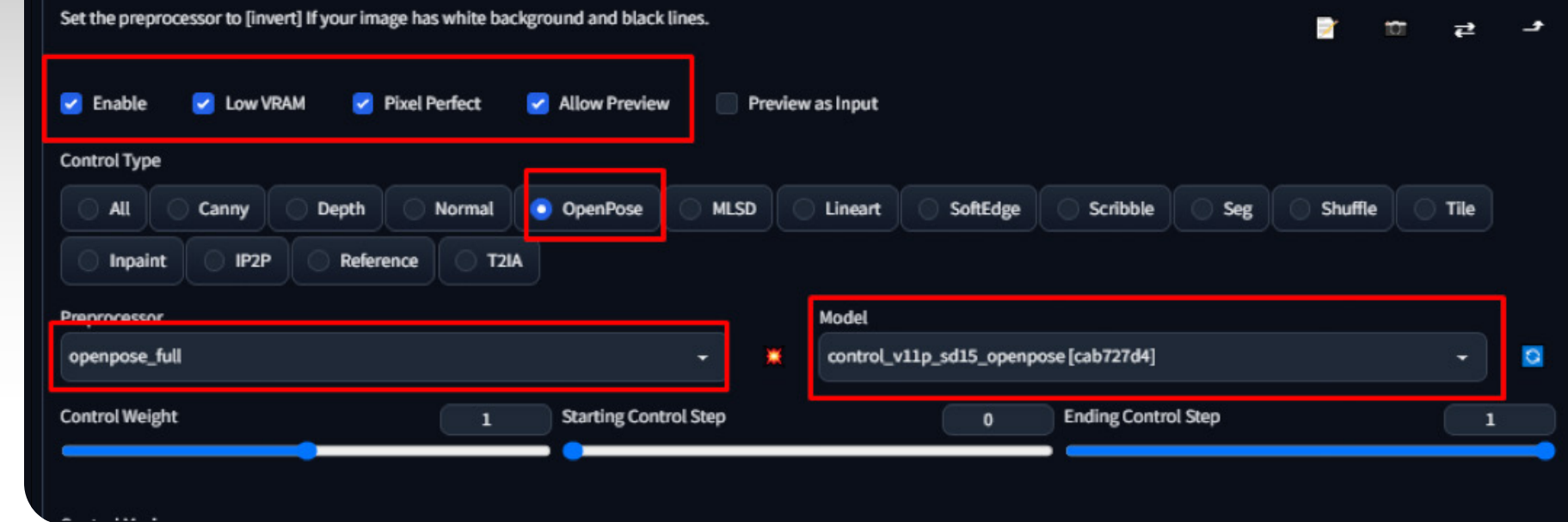
Make sure all the options on your side match this image including the checkboxes and the CN model etc.

Click Generate.

Drag your body pose map, into the CN image reference.

Add a prompt such as: “A muscular woman, grabbing her fist”

This same method can be done with various CN input methods such as Canny Edge etc.



ControlNet Use-Cases

Designing QR Codes

First things first, generate your QR code, you can do this through any QR code generator online, but I used **this one** for this example. Fill in the information, the URL, preferably use a URL-shortener such as bit.ly, then hit “Generate code”



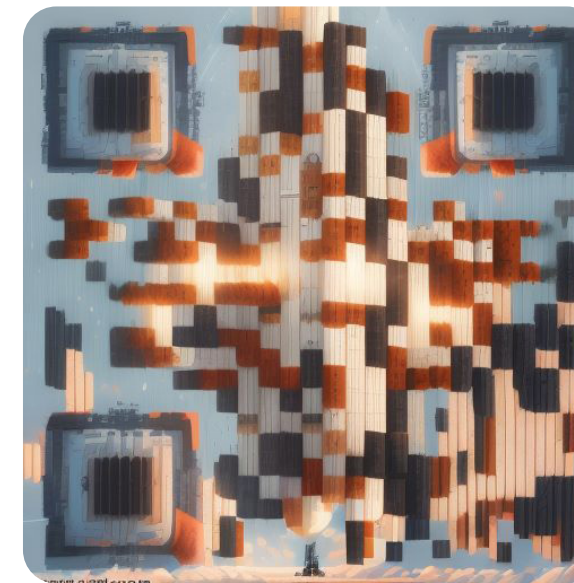
Code Type:	<input checked="" type="radio"/> QR Code (recommended) <input type="radio"/> Data Matrix (only ASCII chars) <input type="radio"/> Aztec Code (only ASCII chars) <input type="radio"/> Micro QR Code (only ASCII chars)
Web Site URL: *	<input type="text" value="https://nabilkhaled.com/"/>
URL Shortening: <small>(will use full url in case of error)</small>	<input type="text" value="bitly"/>
Error Correction Level: <small>(only for QR Code)</small>	<input type="text" value="High"/>
Block Size in Pixels:	<input type="text" value="20"/>
Margin Size in Blocks:	<input type="text" value="1"/>
Output Type:	<input type="text" value="Portable Network Graphics (PNG)"/>
Foreground Color:	<input type="text" value="#000000"/> <input type="checkbox"/> Transparent
Background Color:	<input type="text" value="#FFFFFF"/> <input type="checkbox"/> Transparent
<input type="button" value="Generate Code"/> <input type="button" value="Reset Form"/>	

ControlNet Use-Cases

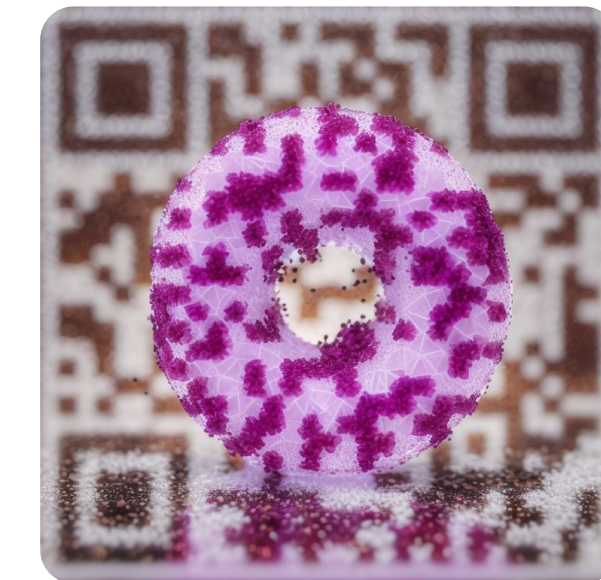
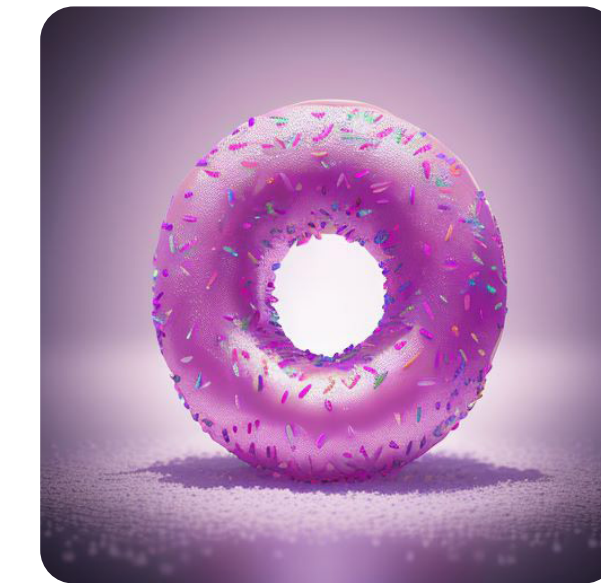
Designing QR Codes

While this technique may not have been entirely perfected yet, it represents a compelling use-case for ControlNet. Here, you can redesign and embellish ordinary QR codes using two ControlNet units concurrently. I've seen some exceptional designs crafted by artists online, and in the ensuing pages, I'll present some examples from my own experiments with this use-case. Following that, I'll guide you on how you can accomplish this by yourself.

Without reference image



With reference image



ControlNet Use-Cases

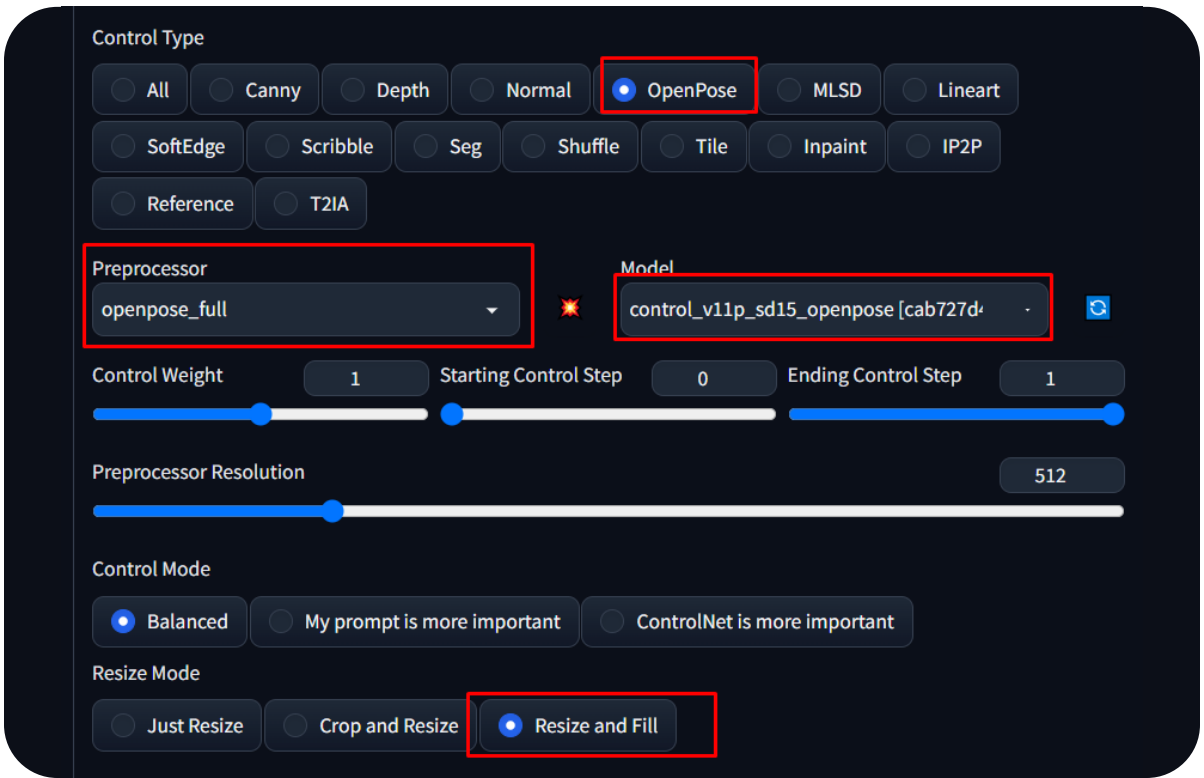
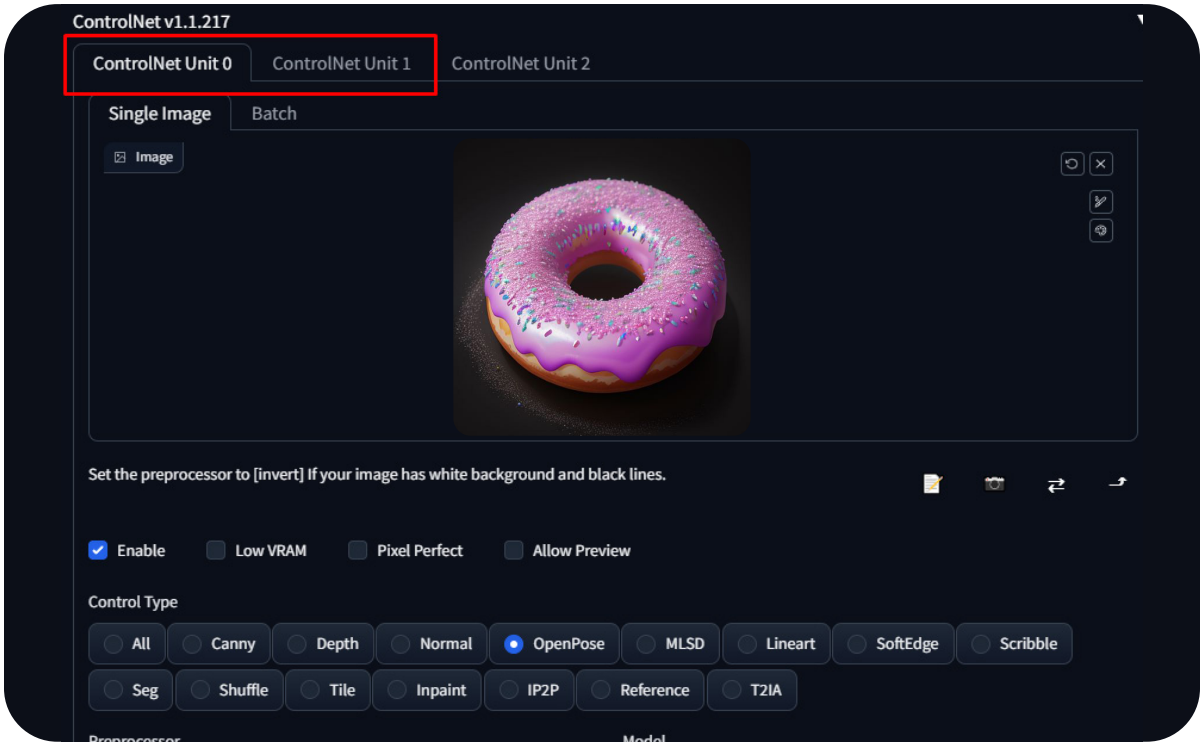
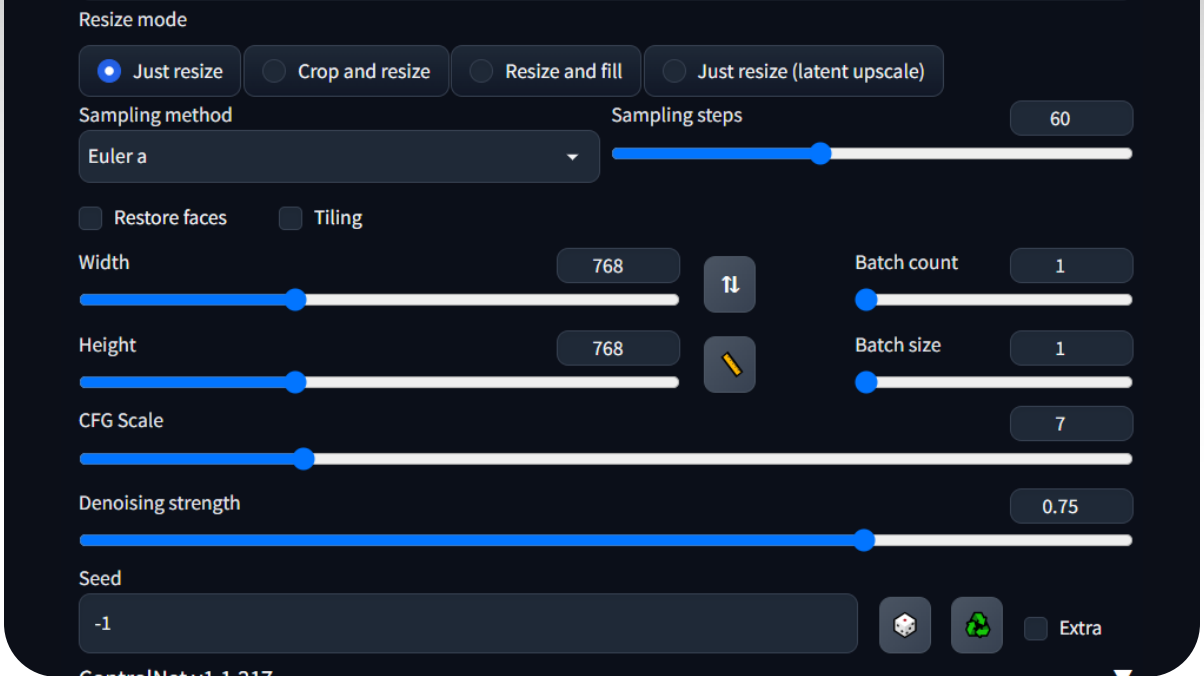
Designing QR Codes

Next, navigate to the 'img2img' tab. While you are free to experiment with various settings and parameters, recommended ones can be found in the image provided here.

For this demonstration, we will be using a reference image of a doughnut, which could be cleverly used for a doughnut shop menu.

Following that, access the ControlNet settings. Here, select 'OpenPose' as your control type. Ensure that you are in the 'Resize and Fill' mode, and the 'enable' checkbox is activated.

Transition to your second ControlNet unit next. In case you don't have multiple ControlNet units activated, I recommend watching [this video](#). It will guide you on how to activate additional units and walk you through the entire QR-code manipulation process using ControlNet.



ControlNet Use-Cases

Designing QR Codes

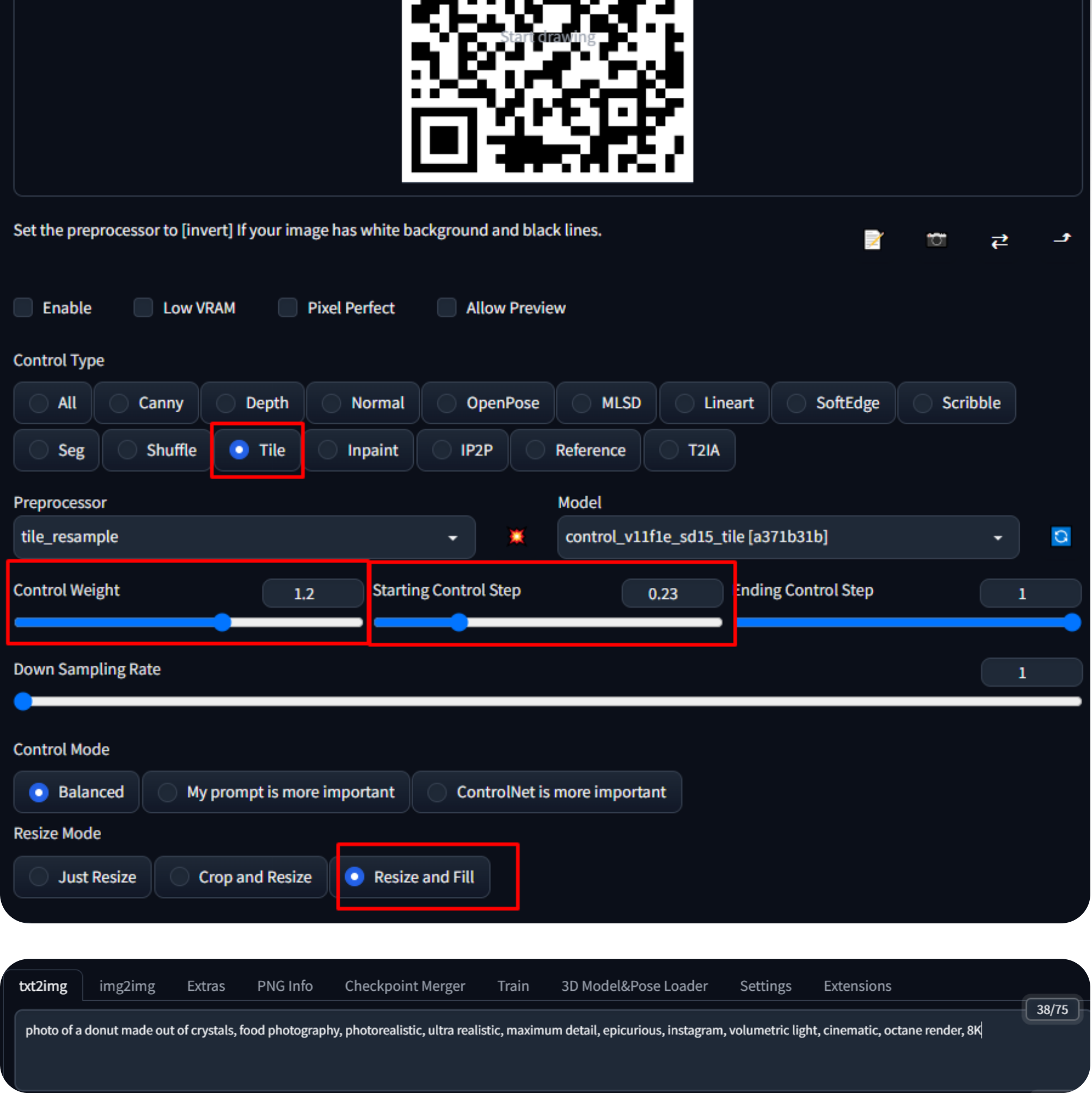
In the second tab, you'll need to adjust your Control Weight to a slightly elevated value, say 1.2, and set the Starting Control Step to around 0.23 or higher. Feel free to adjust these parameters based on your requirements. The values I've used are recommendations [sourced](#) from a YouTube video.

In this tab, you're also required to upload your barcode image and select the 'Tile' Control Type.

Next, insert the original image into the 'img2img' reference box at the top. Enter a descriptive prompt for the existing image and click on 'generate'.

Your initial attempts may result in images that aren't great. But don't hesitate to experiment multiple times until you achieve your desired outcome. The remarkable aspect here is that, despite any potential visual imperfections, the barcodes still function perfectly!

Although this method works quite well, I found another brilliant tutorial [here](#).



ControlNet Use-Cases

There's a lot more happening

The pace at which AI is advancing is truly awe-inspiring, especially when we consider its implications for image generation and the imminent reality of full-fledged video production via Stable Diffusion. Already, it's possible to create flicker-free videos with Stable Diffusion, though this process is quite resource-intensive.

In some exciting developments, individuals have been leveraging 3D modelling tools to create highly customized poses. They've then captured still images from their Blender models and imported them into ControlNet! The process is akin to what I've demonstrated above, but extends to generating multiple frames for video creation, stop-motion artwork, and more!

Here are some resources for additional information:

- [Resource 1](#)
- [Resource 2](#)

Please be advised: as you delve into this subject and explore the work of prominent bloggers who have conducted research and experimentation, you may encounter NSFW content.



Textual Inversion Embedding

Textual inversion embeddings are a technique used to incorporate new concepts or features into a model without altering its underlying architecture. This method allows for injecting novel styles, objects, or visual/sensory elements into generated images using text-to-image models.

In practical terms, this means that fascinating results can be achieved without the need to install modified models. Instead, small files obtained from Stable Diffusion community websites (such as civitai) can be added. It is also possible to create custom textual inversion embeddings.

The versatility extends further as multiple embeddings can be used simultaneously, effectively blending pre-trained styles akin to merging modified models. Notably, these embeddings perform significantly better in Stable Diffusion V1/2.0 compared to V5/1.4.

To activate an embedding, a keyword is required. For instance, to activate the VikingPunk embedding, the keyword "Vikingpunk" must be used. These keywords are typically provided on the same page where the embedding is downloaded.

Using embeddings with keywords is a highly effective and user-friendly method to enhance image quality. However, when using such content for commercial purposes, it is important to be aware of potential constraints. In a dedicated section of this book, we will explore the specific considerations surrounding alternative models and textual inversion embeddings for commercial usage.



Applications

Textual Inversion embedding offers diverse applications:

- Personal Use: Create custom images for profiles, backgrounds, or artwork.
- Marketing and Advertising: Generate unique visuals for campaigns, capturing attention and standing out.
- Educational Materials: Enhance illustrations, diagrams, and interactive exercises for better comprehension.
- Concept Visualization: Transform abstract or complex concepts into visual representations.
- Content Creation and Design: Generate visual assets for websites, blogs, book covers, and artistic compositions.

Considerations for Textual Inversion embedding:

- Quality of text: Clear and concise descriptions yield better results.
- Computational cost: GPU usage and batch size may impact performance.
- Not infallible: While effective, perfect image generation is not guaranteed.
- In summary, Textual Inversion embedding is a powerful tool for creating realistic and engaging images from text descriptions.

Understanding its limitations is crucial before implementation.



How to Install:

Prominent users in the Stable Diffusion community, such as ShadowXShinigami and ConflictX, have developed groundbreaking models like the midjourney papercut and CG animation models. These models cater to specific styles, including Disney and Viking punk. With Stable Diffusion, you can easily filter and incorporate textual inversion models into your workflow.

Just ensure that the embeddings are trained on the version of the model you're using (e.g., V2.1). By downloading the PT or PNG files and placing them in the designated "Embeddings" folder, you can seamlessly access and utilize the loaded embeddings within Stable Diffusion. (Screenshot example from [AIPreneur](#)). Typically, this is the address for your embeddings folder unless you chose a different path during the installation and setup phase: `C:\Users\{name}\stable-diffusion-webui\embeddings`

Once you run your webui, you should see all your newly installed embeddings being run as you can see in the image here.

PC > Local Disk (C:) > Users > [redacted] > stable-diffusion-webui > embeddings

Name	Date modified	Type	Size
anthro.pt	27/05/2023 6:34 am	PT File	41 KB
CGI_Animation.png	27/05/2023 6:34 am	PNG File	603 KB
CGI_Animation.pt	27/05/2023 6:34 am	PT File	17 KB
CGI_Animation-185.png	27/05/2023 6:34 am	PNG File	658 KB
CGI_Animation-185.pt	27/05/2023 6:34 am	PT File	17 KB
CGI_Animation-245.png	27/05/2023 6:34 am	PNG File	621 KB
CGI_Animation-245.pt	27/05/2023 6:34 am	PT File	17 KB
cruise_ship_on_wave_kc16-v3-6250.png	27/05/2023 6:34 am	PNG File	1,196 KB
kc16-v3-6250.pt	27/05/2023 6:34 am	PT File	65 KB
kc16-v4-5000.pt	27/05/2023 6:34 am	PT File	65 KB
mdjrny-ppc.pt	27/05/2023 6:34 am	PT File	33 KB
mdjrny-ppc-150.png	27/05/2023 6:34 am	PNG File	949 KB
midjourney.pt	27/05/2023 6:35 am	PT File	5 KB
Place Textual Inversion embeddings here....	26/03/2023 3:56 am	Text Document	0 KB
README.md	27/05/2023 6:34 am	Markdown Source...	2 KB
remix.pt	27/05/2023 6:34 am	PT File	41 KB
VikingPunk.pt	27/05/2023 6:34 am	PT File	25 KB
VikingPunk-63.png	27/05/2023 6:34 am	PNG File	864 KB
vray-render.pt	27/05/2023 6:34 am	PT File	25 KB
vray-render-500.png	27/05/2023 6:34 am	PNG File	845 KB

```
C:\WINDOWS\system32\cmd.exe
Already up to date.
venv "A:\AI\SUPER SD 2.0\stable-diffusion-webui\venv\Scripts\Python.exe"
Python 3.10.5 (tags/v3.10.5:f377153, Jun 6 2022, 16:14:13) [MSC v.1929 64 bit (AMD64)]
Commit hash: 685f9631b56ff8bd43bce24ff5ce0f9a0e9af490
Installing requirements for Web UI

Launching Web UI with arguments: --deepdanbooru --xformers --api --cors-allow-origins=https://www.painthua.com
Loading config from: A:\AI\SUPER SD 2.0\stable-diffusion-webui\models\Stable-diffusion\v2-1_768-nonema-pruned.yaml
LatentDiffusion: Running in v-prediction mode
DiffusionWrapper has 865.91 M params.
Loading weights [e1542d5a] from A:\AI\SUPER SD 2.0\stable-diffusion-webui\models\Stable-diffusion\v2-1_768-nonema-pruned.ckpt
Applying xformers cross attention optimization.
Model loaded.
Loaded a total of 31 textual inversion embeddings
Embeddings: anthro, by remix, cgi_animation, CinemaHelper, EMB_MJ-100, EMB_MJ-1000, EMB_MJ-1100, EMB_MJ-1200, EMB_MJ-200, EMB_MJ-300, EMB_MJ-400, EMB_MJ-500, EMB_MJ-600, EMB_MJ-700, EMB_MJ-800, EMB_MJ-900, EMB_MJ, greg rutkowski, InkPunk768, knollingcase, kuvshinov, midjourney, painted_landscape, PhotoHelper, sakimi-style, SDA768, VikingPunk-63, VikingPunk, VintageHelper, vray-render, _SamDoesArt2_
Running on local URL: http://127.0.0.1:7860

To create a public link, set 'share=True' in 'launch()'.
```


Examples

In this case, I wrote a relatively simple prompt

To trigger the TIV, I just had to add the specific embedding's keyword "VikingPunk" in my prompt.

Now this is quite a simple prompt that could have been highly refined using better keyword order, terminology, punctuation etc.

But you can clearly see how the embedding made a huge difference. Using the standard Automatic1111 Stable Diffusion V2.1 model, almost as if you were using a modified model!

Disclaimer: These images are prompted by the author, but were inspired by very similar prompts from the community and the creator of this model

photo of a (police car:1.1), siren_lights, city, (apocalyptic_scene:1.4), (cyberpunk style:1.3), vibrant colors, depth of field, (cannon R5, f1.4, 24mm lens:1.3), **VikingPunk**, detailed, (superrealistic:1.4), 8k, 4k, depth of field, (hdr:1.2)



Without



With

Examples

Here you can see a few other examples I tried out to illustrate the idea further.

Note: all of these images were done using the default Stable Diffusion V2.1 model, after injecting a different textual embed for each prompt.

Note: Similar results can be achieved with customized models or LORAs

Some sources for embeddings that you can use in Stable Diffusion can be found [here](#):

Source



Source



Multiple Sources



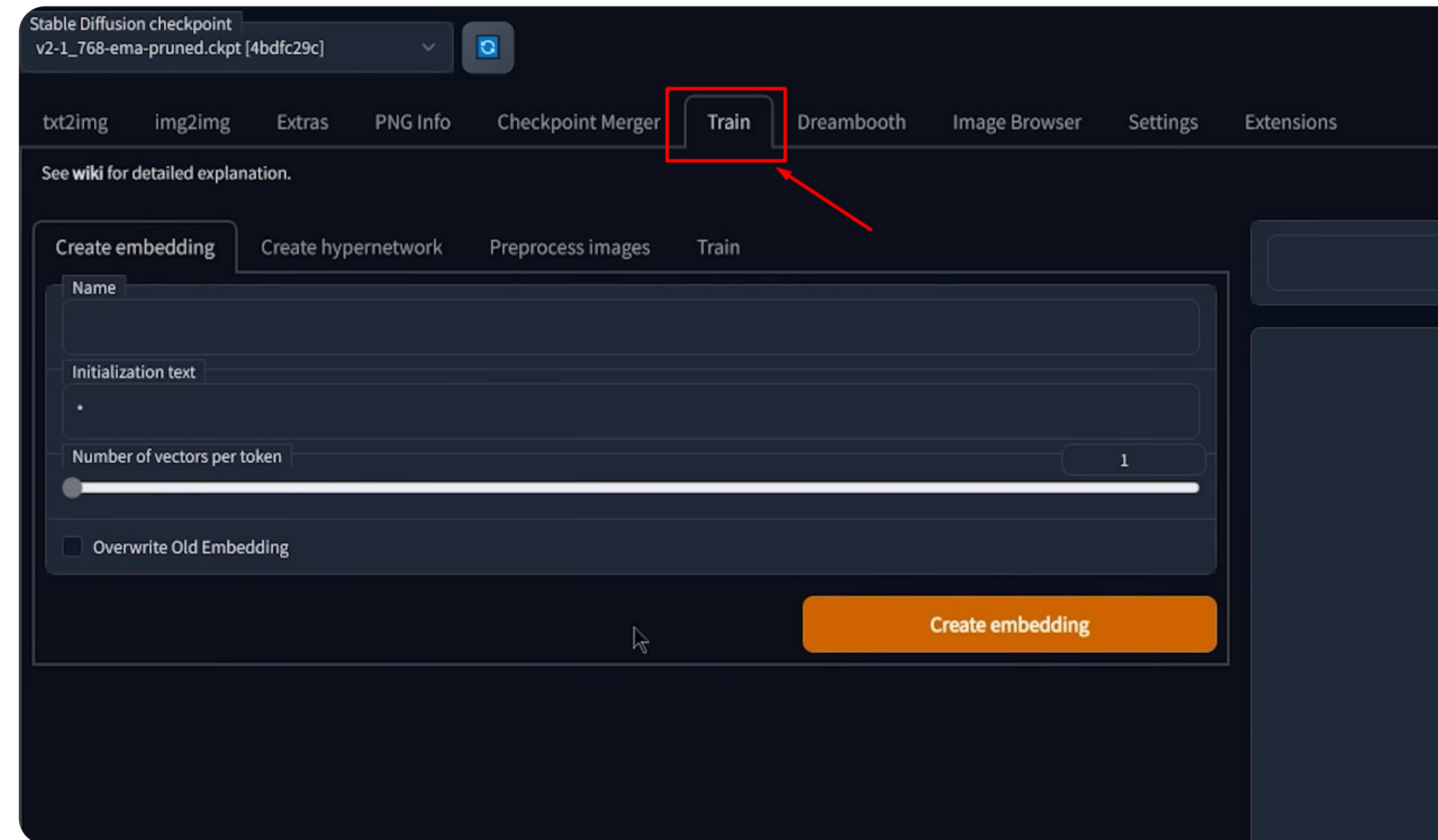
Disclaimer: These images are prompted by the author, but were inspired by very similar prompts from the community and the creator of this model

Create Your Own

In this book, we will provide a brief overview of creating and training textual inversion embeddings, but for more detailed instructions, there are various online sources available. Towards the end of this book, we will share some of these sources with you.

To begin, take a look at the "train" tab in your GUI. It is a useful feature that I initially overlooked until I started exploring embeddings and training my own.

For further guidance, you can refer to [Aitrepreneur's YouTube video](#) on the topic. This resource will provide you with more in-depth information and insights.



Create Your Own

- On a basic level, here are the steps involved in creating a textual inversion embedding:
- Choose a unique and memorable name for your embedding. A format like "NEW-Embed-(NAME)" is recommended to ensure uniqueness and ease of identification.
- Set the number of vectors per token to "8," as suggested by [Aitrepneur](#).
- Navigate to the "Preprocess images" tab in the GUI.
- Use the "Source directory" field to specify the folder where your images are located.
- Enable the options "Create flipped copies," "Auto focal point crop," and "Use Blip for Caption" if desired.
- "Create flipped copies" duplicates the images by creating flipped versions of each file, effectively doubling the number of images.
- "Auto focal point crop" automatically crops the images while keeping the subject in the center.
- "Use Blip for Caption" scans the images and generates a text file that describes the content of each image, similar to alt text used in social media or websites.

By following these steps, you can begin the process of creating your textual inversion embedding.

See wiki for detailed explanation.

Create embedding Create hypernetwork Preprocess images Train

Name

Initialization text

Number of vectors per token

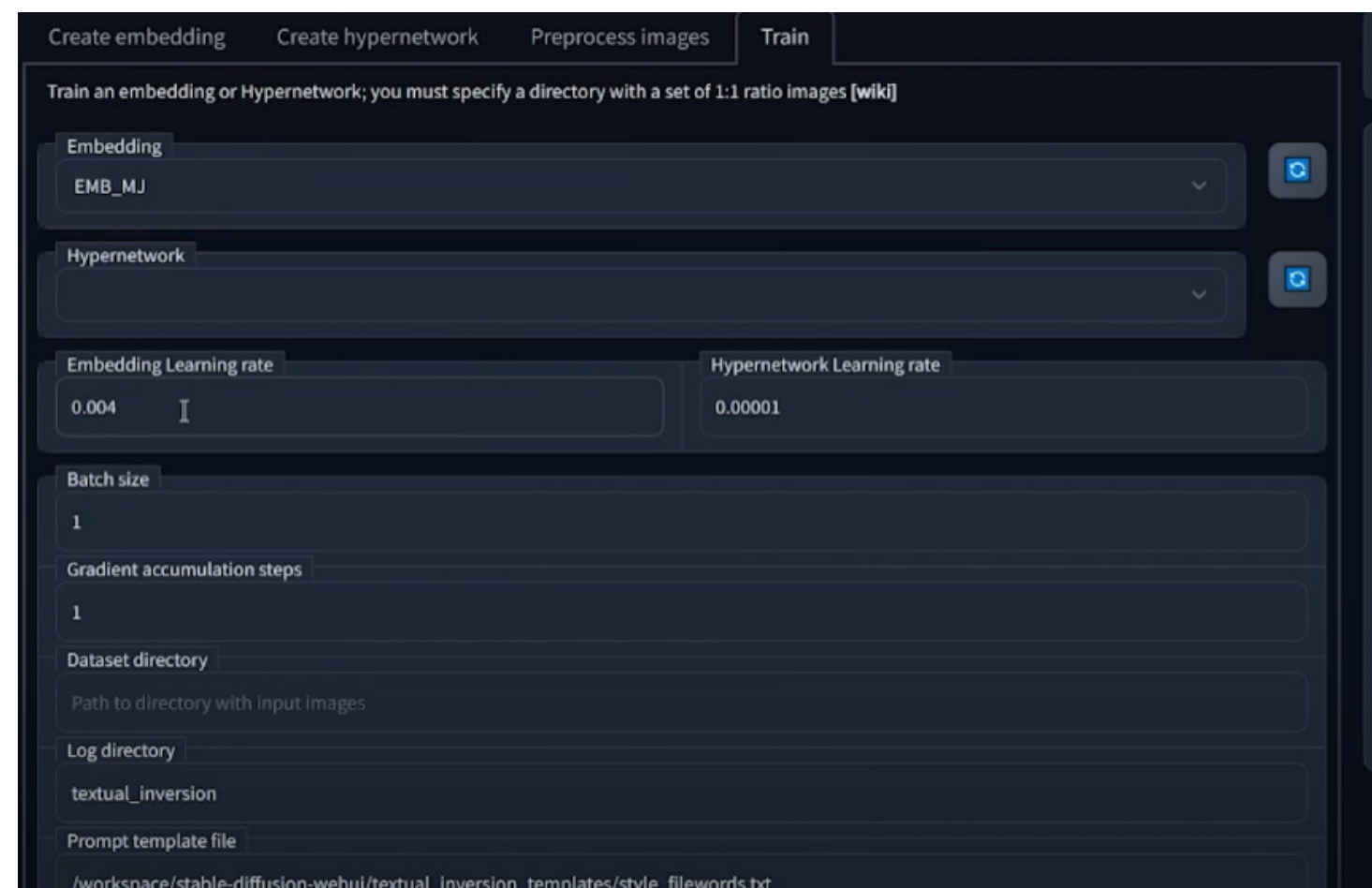
☐ Overwrite Old Embedding

Create

Additional Recommended Settings from AiPreneur

Under "Gradient accumulation steps," it is recommended to set the value to half the number of total images you plan to use for training the model. This helps optimize the training process.

Remember to paste your folder directory here, the folder created from the "preprocessing step"



The screenshot shows the 'Train' tab of the AiPreneur interface. It includes fields for 'Embedding' (set to EMB_MJ), 'Hypernetwork', 'Embedding Learning rate' (0.004), and 'Hypernetwork Learning rate' (0.00001). Below these are sections for 'Batch size' (1), 'Gradient accumulation steps' (1), 'Dataset directory' (Path to directory with input images), 'Log directory' (textual_inversion), and 'Prompt template file' (/workspace/stable-diffusion-webui/textual_inversion_templates/style_filewords.txt).



This close-up shows the 'Dataset directory' field with the value 'Path to directory with input images'. Two red arrows point to this field, highlighting its importance. Below it, the 'Log directory' is set to 'textual_inversion' and the 'Prompt template file' is set to '/workspace/stable-diffusion-webui/textual_inversion_templates/style_filewords.txt'.



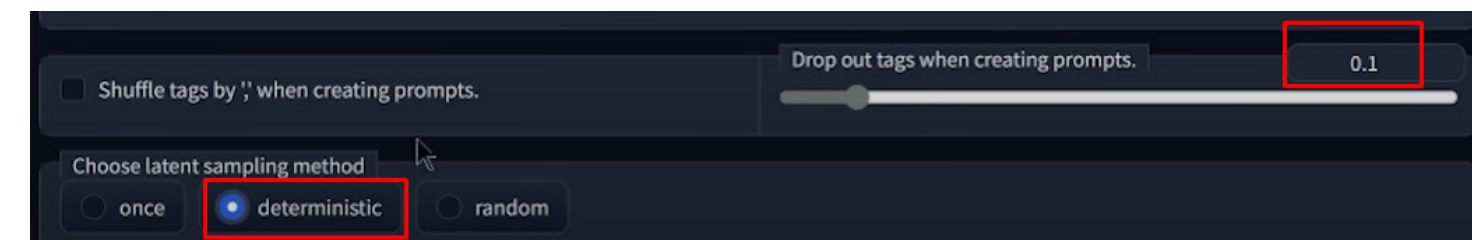
Additional Recommended Settings from AiPreneur

In the "txt2img" tab, there are a few important steps to follow:

Specify your desired prompt for "preview" images that the model will generate every 50 steps. Include the name of your embedding at the end of the prompt. For example:

"Photo of a muscular man, gazing at the sky, natural landscape, centered subject, uplit, natural light, radiant colors, {NAME OF EMBED}"
Adjust the dropout tags slider to a value of 0.1.

Select the latent sampling method as "deterministic" for generating consistent results.

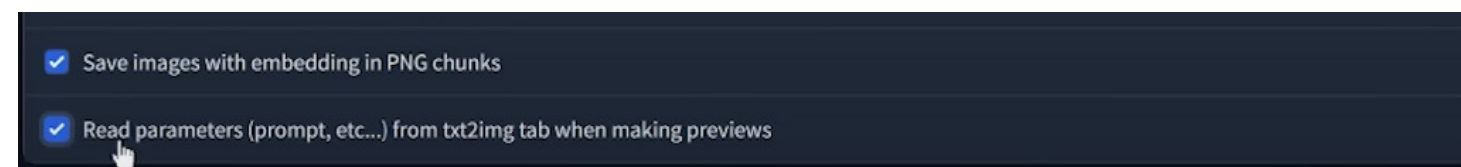


Recommended values for these variables:

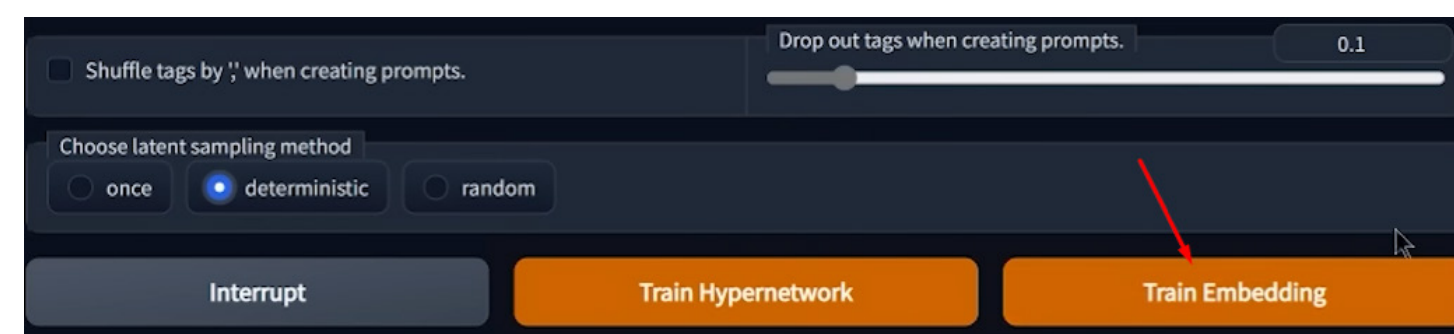


Additional Recommended Settings from AiPreneur

Check these two boxes



Next go ahead and click on Train embedding



After the training process is complete, you can access your generated textual embeddings by following these steps:

- Navigate to the textual embeddings folder in your Stable Diffusion web UI folder.
- Inside the textual embeddings folder, you will find a folder named with the current date.
- Within this date-specific folder, you will find a subfolder with the name of your embedding.
- Inside the embedding subfolder, there will be a folder named "embeddings".
- In the "embeddings" folder, you will find your .pt files, which are your generated textual inversion embeddings.

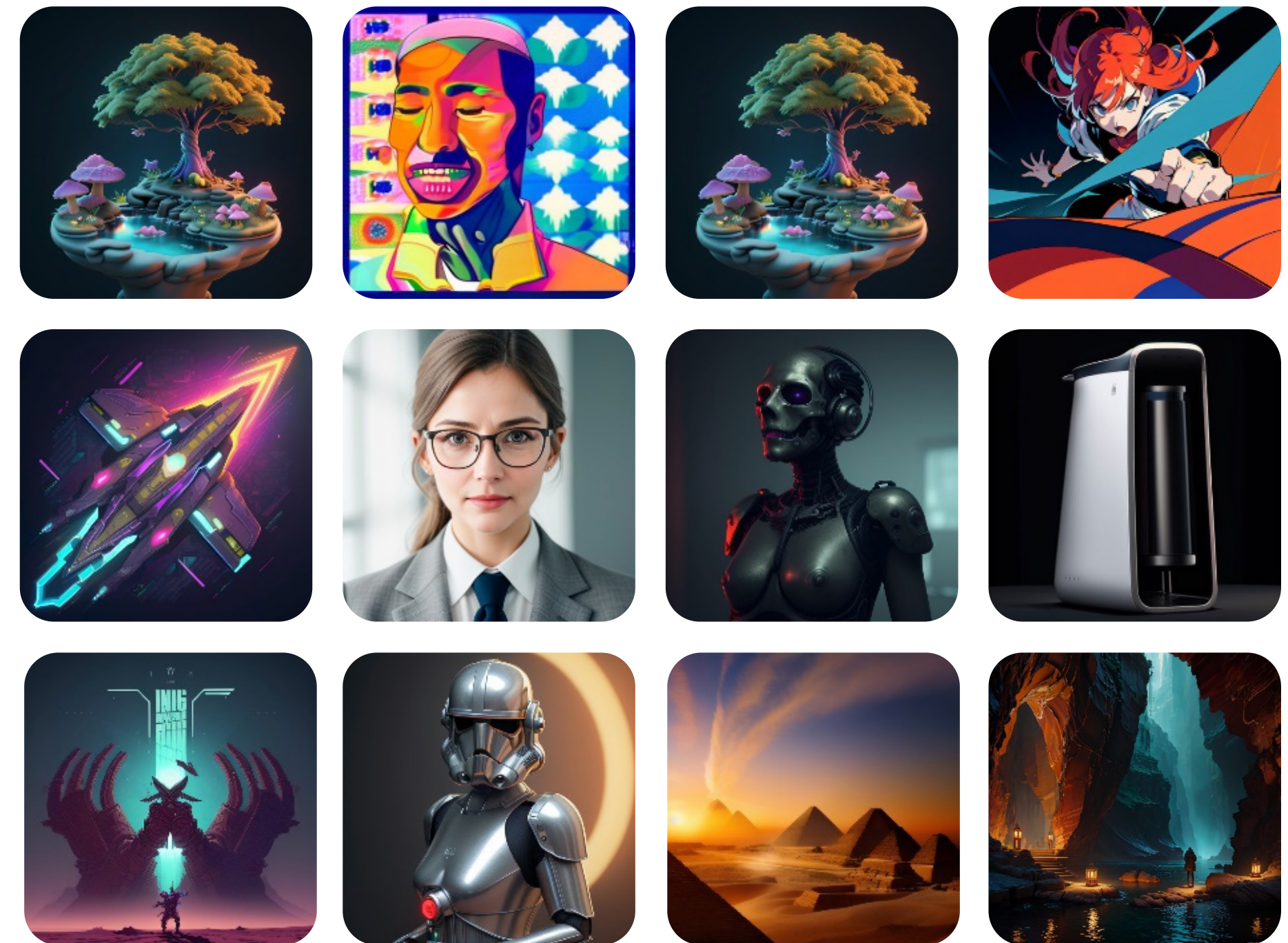
These .pt files represent the embeddings you have created, similar to the ones you downloaded from other users in the community. Now, you have your own customized embeddings to use for generating images.



Marketing Applications

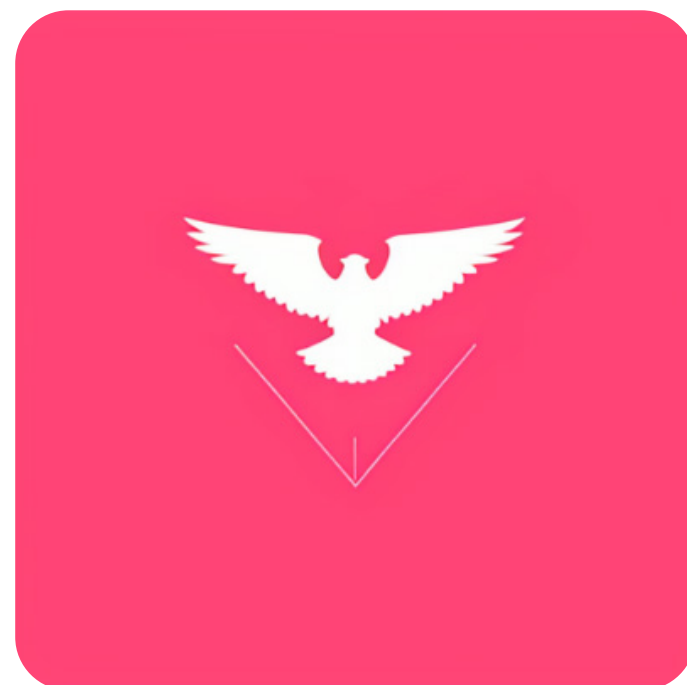
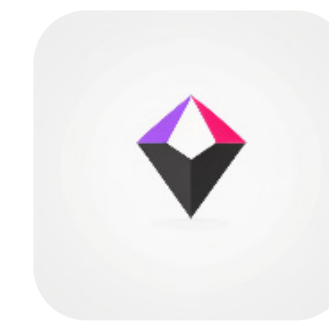
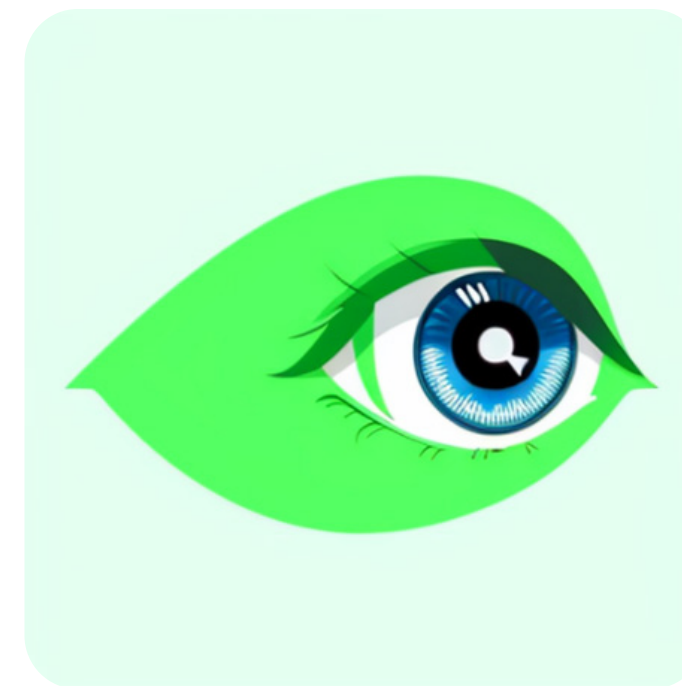
By now you can imagine how versatile and diverse the control you can have via Stable Diffusion is.

The examples shared are relatively obvious and simple to learn use cases in marketing, but if you really think about it, the possibilities are endless



Marketing Applications

Logos & Emblems

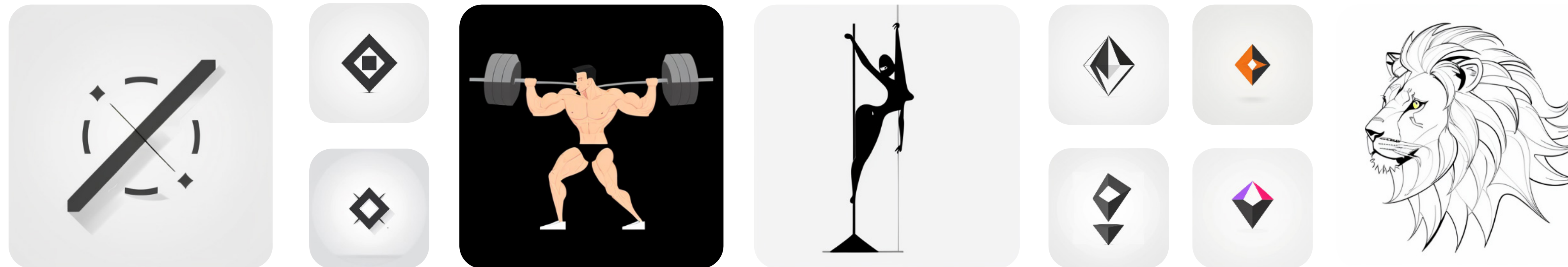
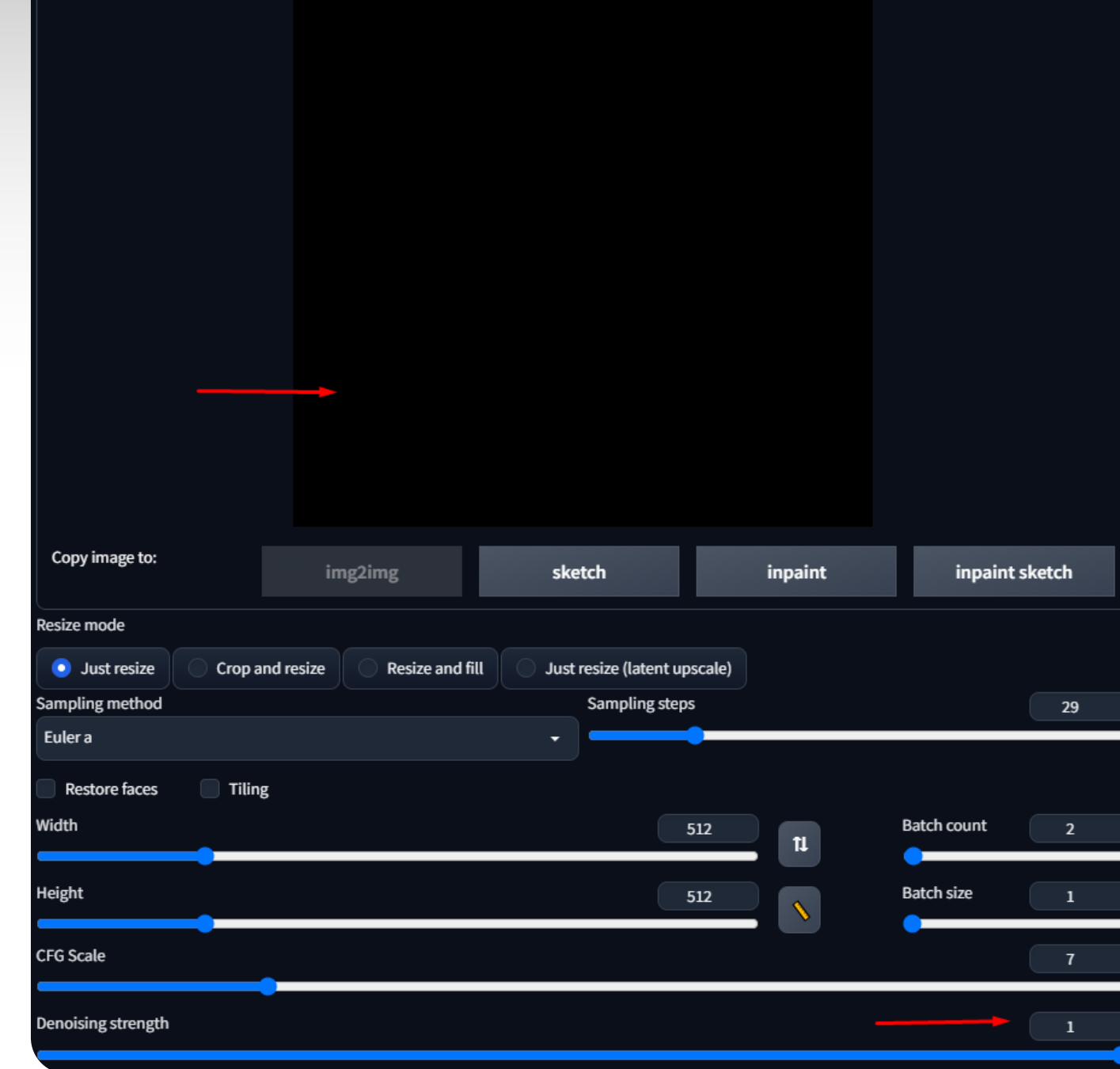


Marketing Applications

Logos & Emblems

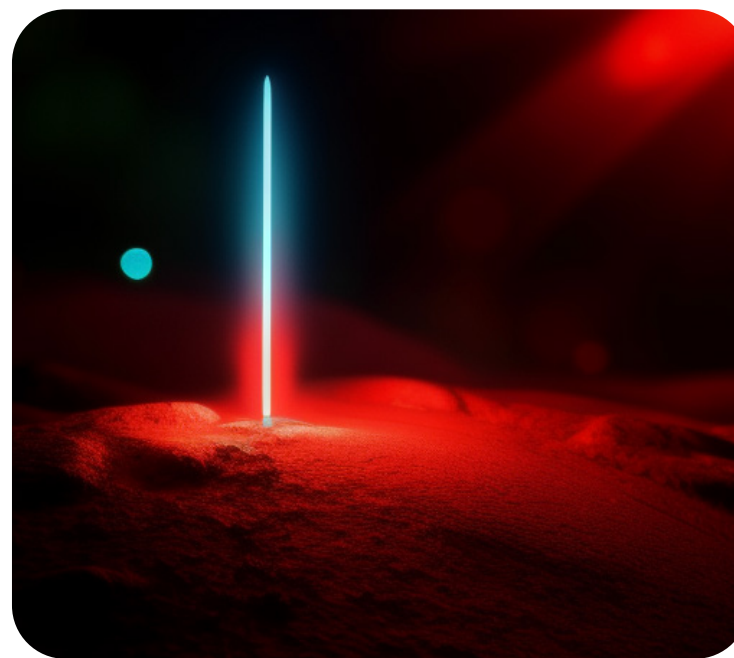
Pro Tip:

Start by selecting a blank monochrome image, preferably in black and white, in your "img2img" tab. Subsequently, adjust your denoising strength to level 1 for optimal results. When entering prompts, make sure they are simple and concise. Experiment with various sampling methods and different fine-tuning models to diversify your results. By doing so, you might be surprised by the impressive results you can create.



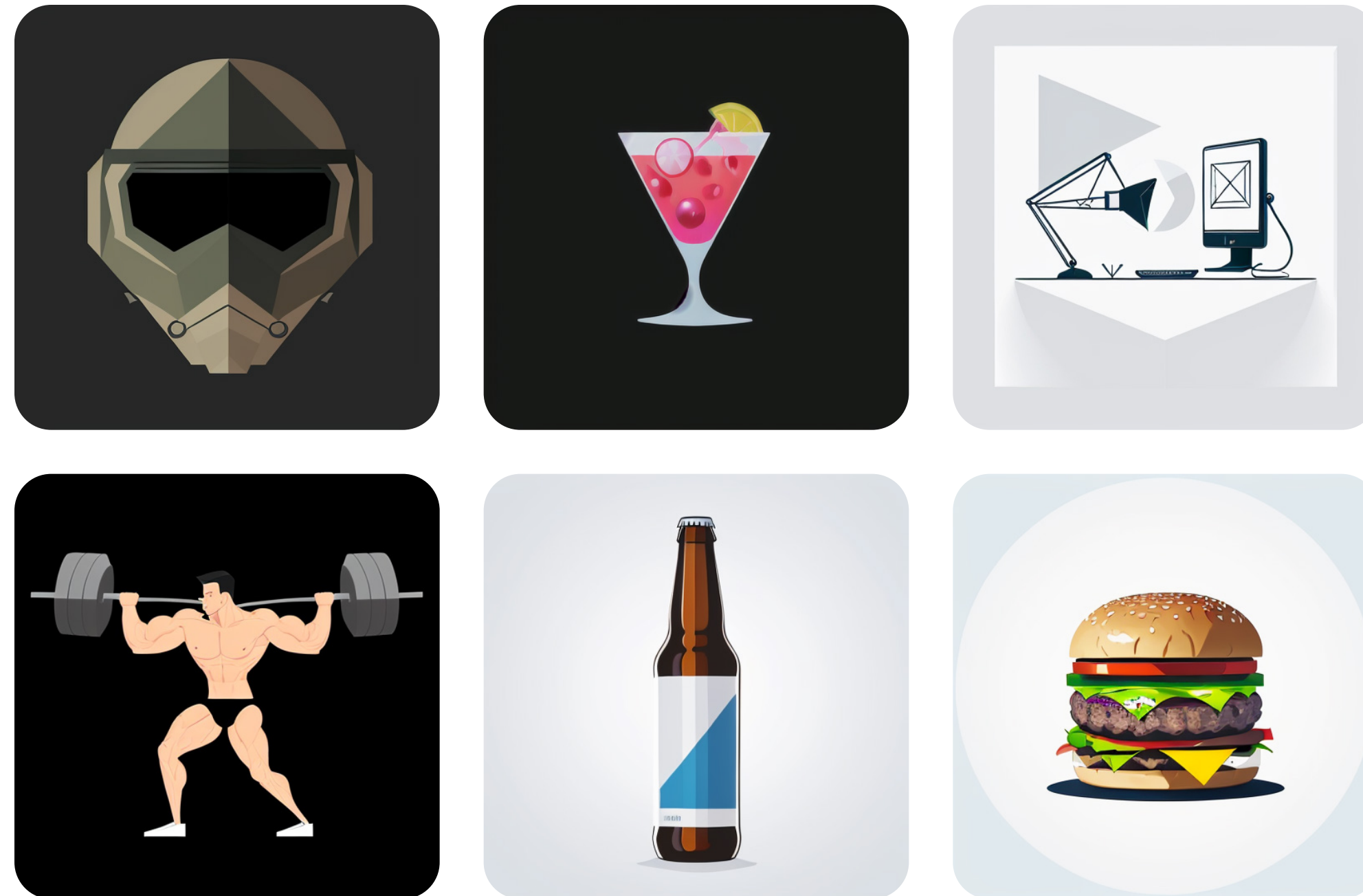
Marketing Applications

Pre-Production: Mood Boards



Marketing Applications

Vector elements



Marketing Applications

Virtual Avatars

As AI, bots, and NLP-based virtual assistants continually evolve, the creation of synthetic avatars for your AI chatbot service has become an effortless process. The demand for such human-like yet unreal avatars is increasingly prevalent across a variety of applications. For instance, they help enhance user experience and maintain a consistent brand image in chatbots and customer support services. They offer a way to preserve privacy when representing individuals in the digital realm. Also, they offer unique character designs in the entertainment and gaming industry, add realism in training simulations, and inspire new art and design avenues. As such, these avatars have become a significant element in today's AI-centric world.



Marketing Applications

3D Conceptualizing

Leveraging AI image-generation models offers a powerful tool in conceptualizing 3D designs and product designs. These models have the capacity to translate abstract ideas into visually stunning and, at times, highly accurate design concepts. Using finetuned or pre-trained models can also help.

In this example, I used the Checkpoint “productDesign_eddiemauro20”
But there are also brilliant LORAs that can compliment these models such as;
[LORA: Product Design Minimalism](#)
[LORA: Product Design Elegant Minimalism](#)



Marketing Applications

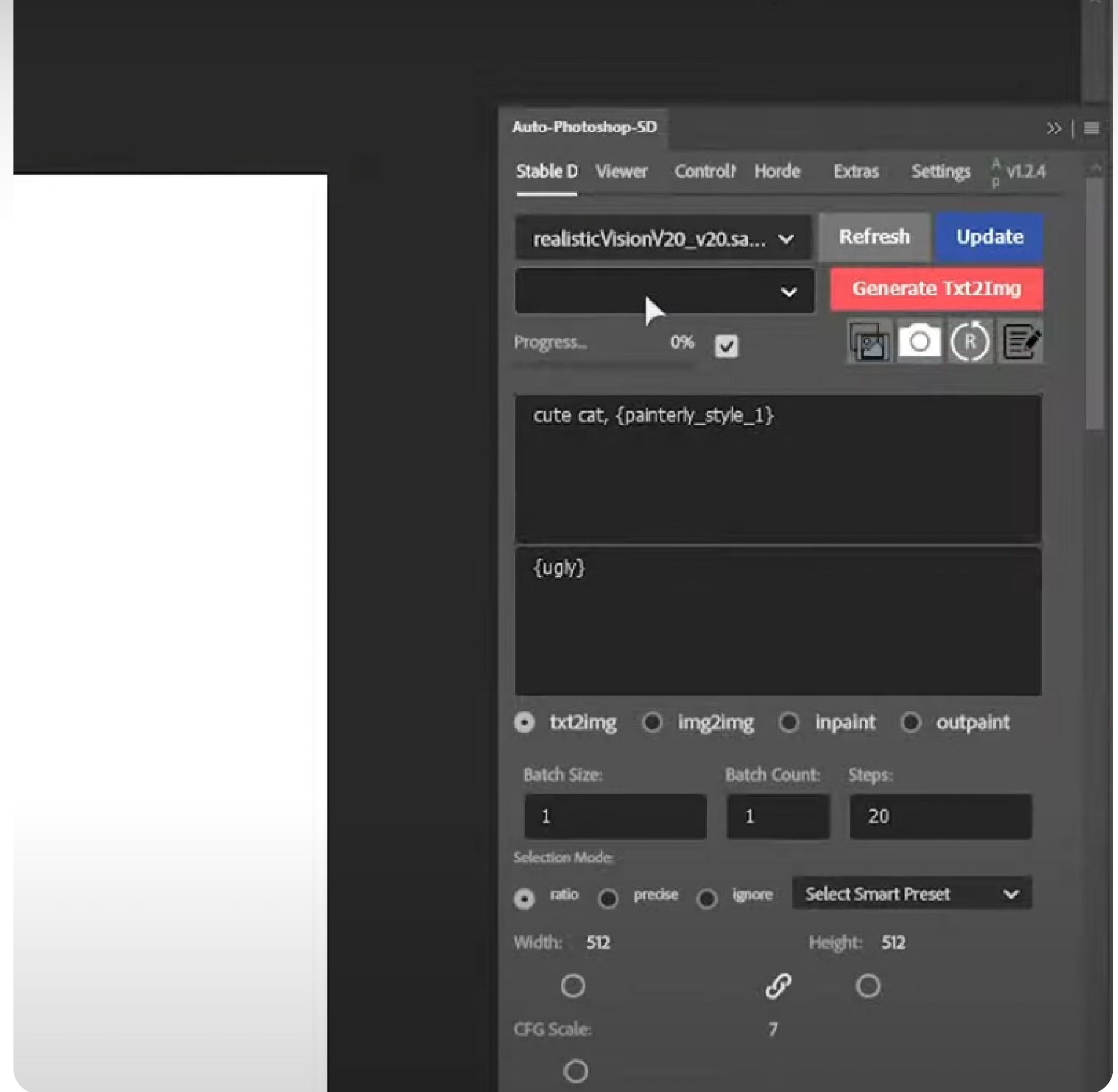
Stable Diffusion: Photoshop plugin

Installation:

First head over to this [link](#). And download the file [Auto.PhotoShop.SD.plugin_v1.2.5.ccx](#)

Double click this file and you will be prompted to install it in your Adobe CC environment.

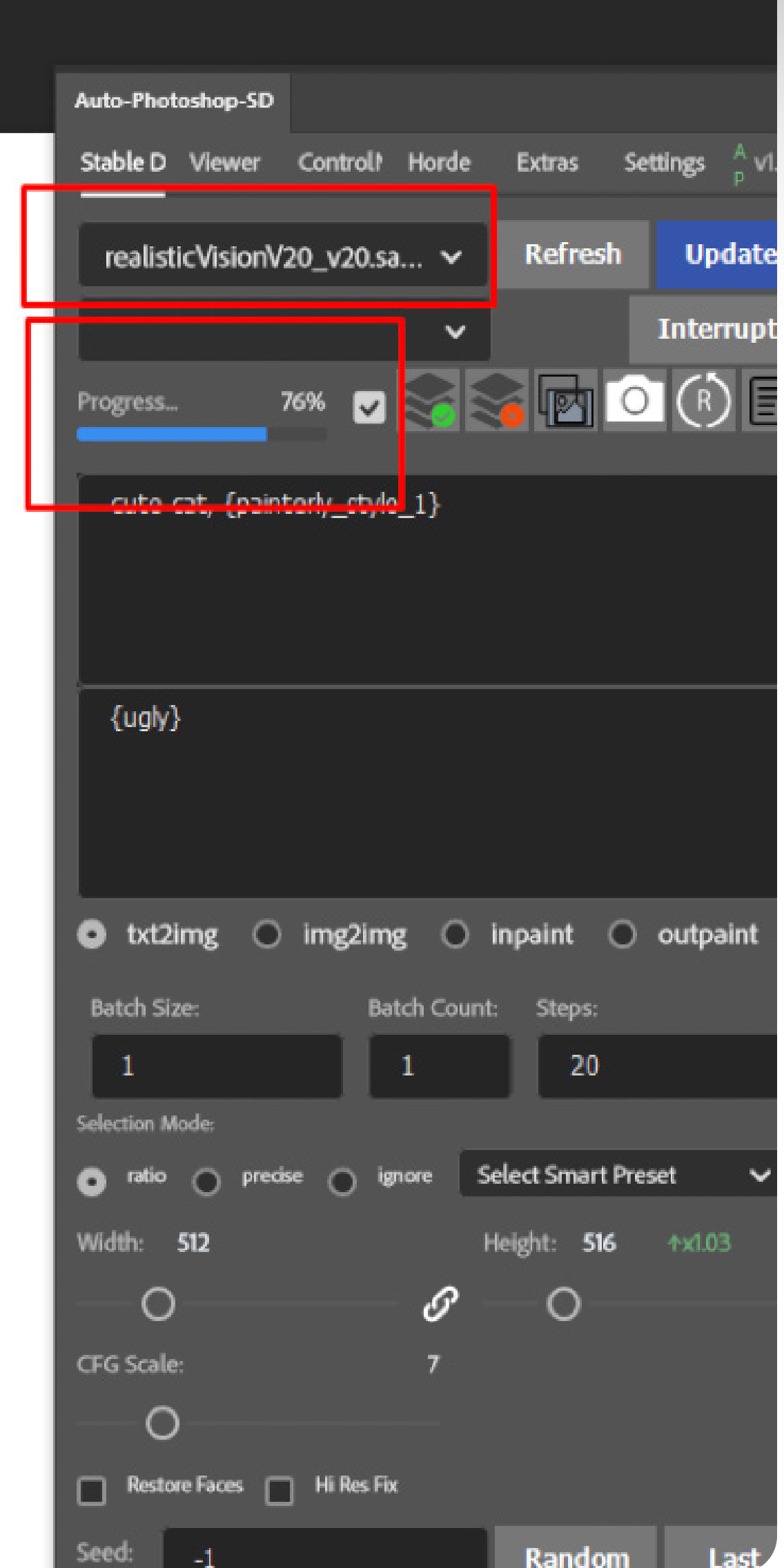
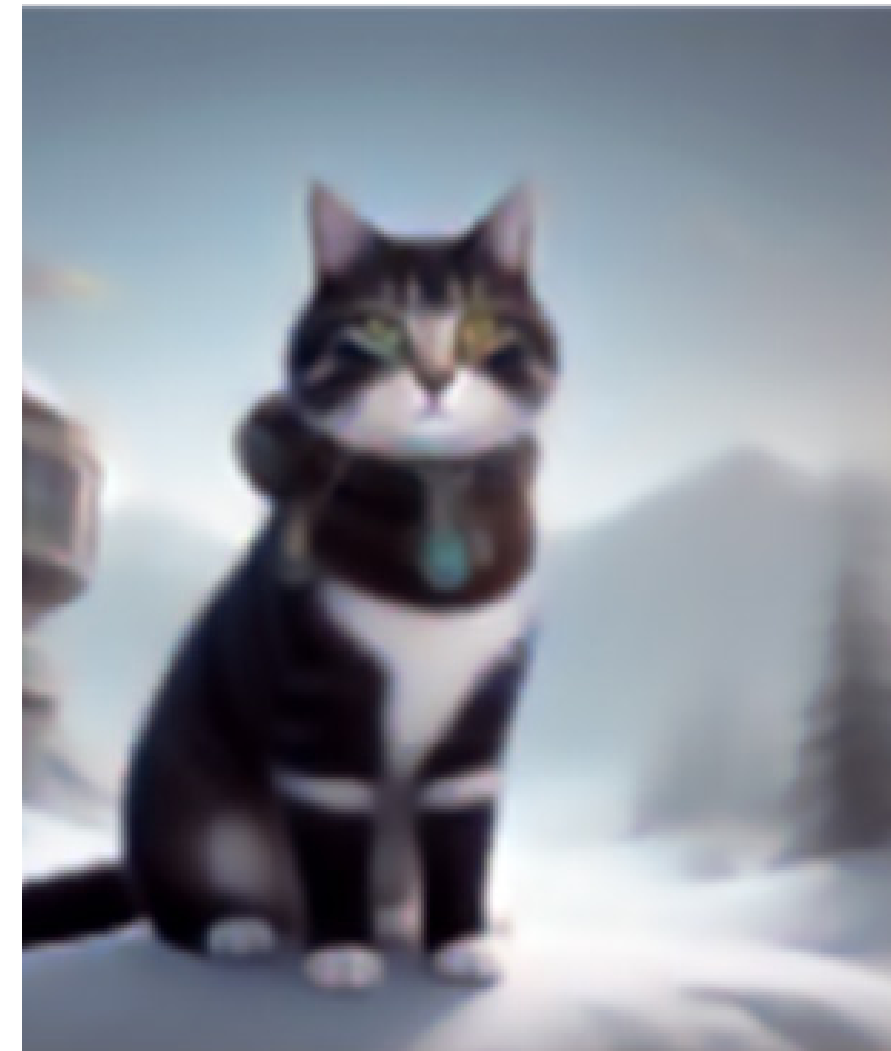
Once it is installed, launch photoshop and head to your Plugins panel where you will find your stable diffusion plugin



Marketing Applications

Stable Diffusion: Photoshop plugin

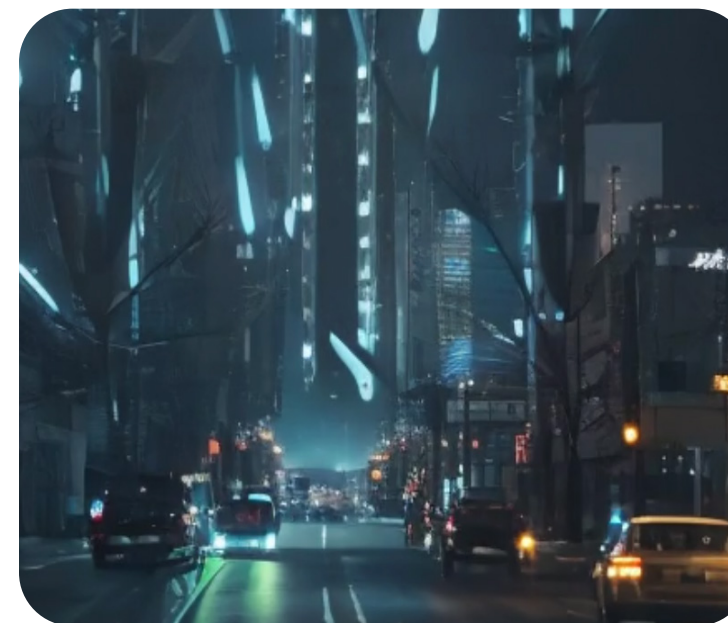
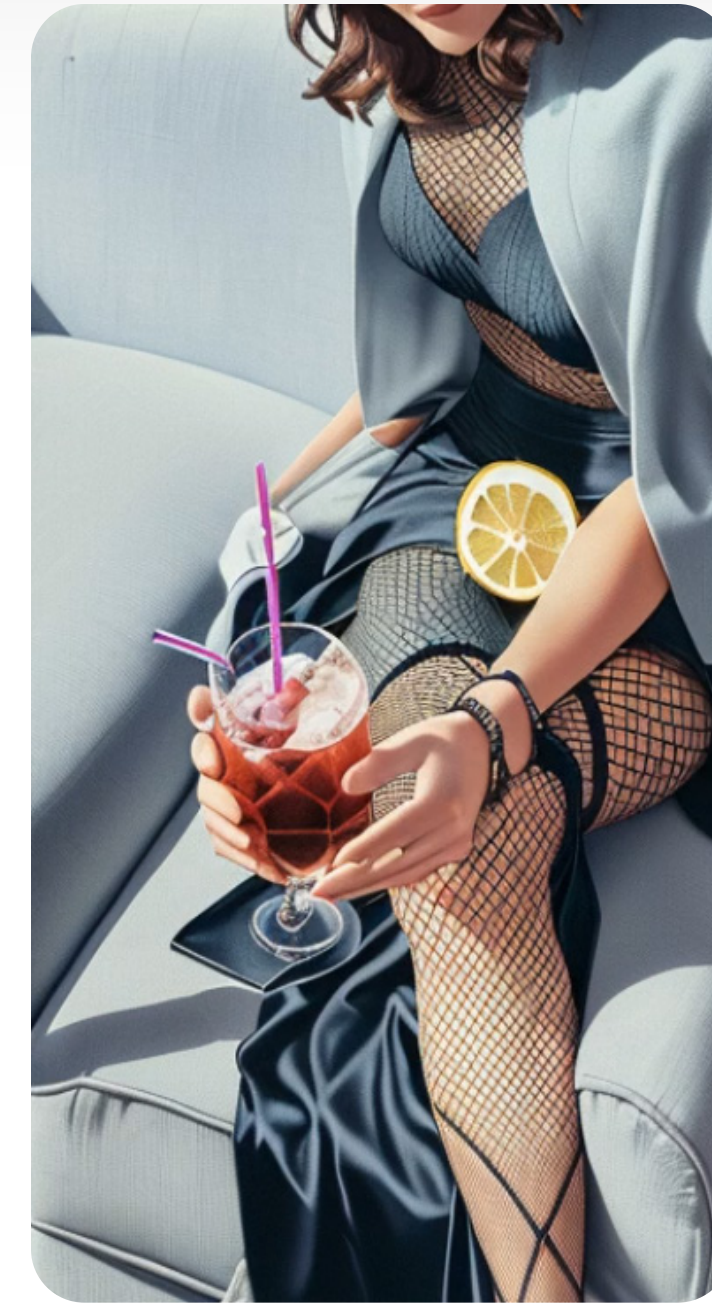
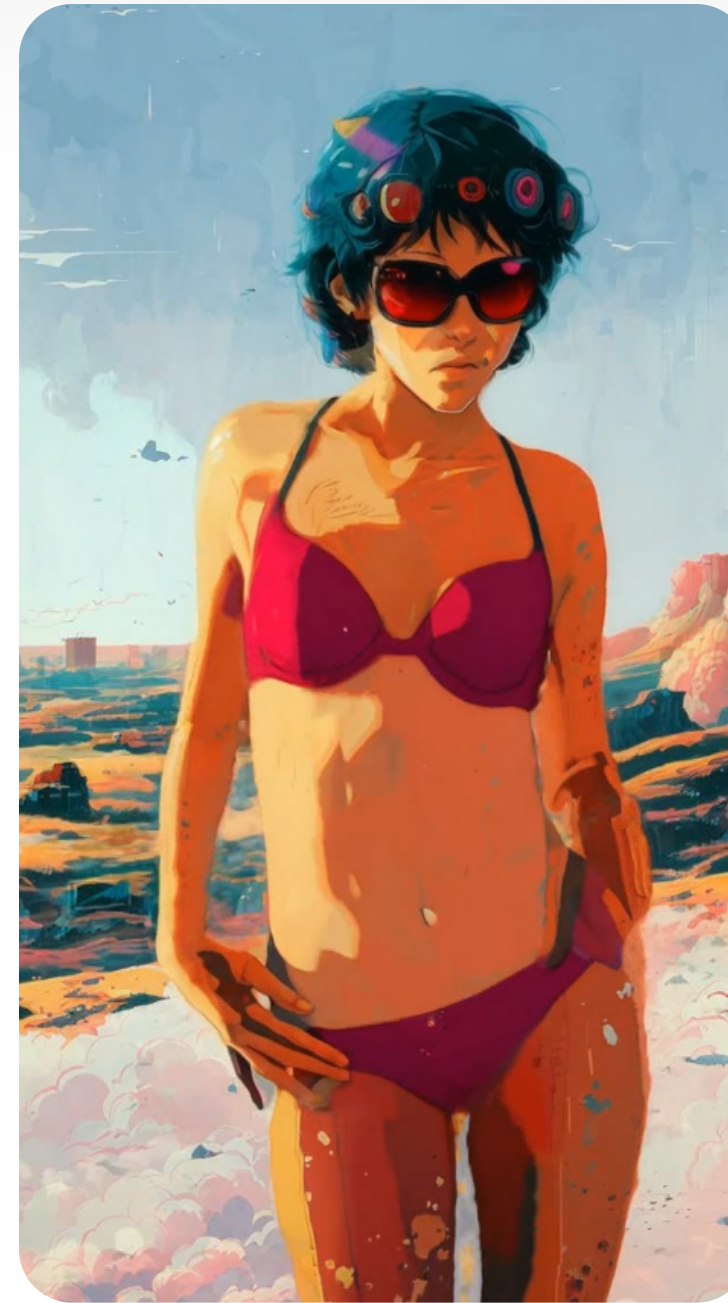
Once you're inside the Plugin environment, you will find that it's very similar to the SD GUI in many aspects, you can choose the model you want to generate images with. The one tip I recommend at the start is that you will need to use your



Marketing Applications

Videography

Since the beginning of Stable Diffusion, many companies have been striving to achieve new heights of image and video generation. Most recently Runway ML launched its cutting-edge text-to-video generator GEN-2, while you also have some very user-friendly tools such as **Kaiber**, **Eluna** and **Genmo**. These new tools are causing a revolutionary shift in videography and animation and it's just the beginning, you have probably already seen many of the millions of videos all over social media where users transform themselves into abstract art, clay, marble or other various visuals using these tools!





05 | SHOWCASING

Showcase Prompts

Many of the prompts I use originate from the inspiring communities at Civit, HuggingFace, Lexica, and GitHub. While I do customize these prompts to fit specific models, the original credit is due to the incredibly inventive Synthetic Data (SD) community. Their creativity and innovation significantly contribute to the work we do.



a photo of a happy couple, sitting on the couch, watching TV, bokeh lighting

Sampler: Euler a
Model: disneyPixarCartoon_v10



a demon made with black stone, Giovanni Strazza sculpt style , bokes sharp focus depth, shadows, mist, fog, very High detailed, cinematic lighting, Cinematic, high detailed, ultra detailed, Accent Lighting, very god colors, realistic, 8k, HDR

Sampler: DPM++ 2M Karras
Model: lyriel_v16



Indie game art,(a photo of a happy couple, sitting on the couch, watching TV, bokeh lighting, detailed skin texture, (blush:0.5), (goosebumps:0.5), subsurface scattering), (Vector Art, Borderlands style, Arcane style, Cartoon style), Line art, Distinct features, Hand drawn, Technical illustration, Graphic design, Vector graphics, High contrast, Precision artwork, Linear compositions, Scalable artwork, Digital art, cinematic sensual, Sharp focus, humorous illustration, big depth of field, Masterpiece, trending on artstation, Vivid colors, trending on ArtStation, trending on CGSociety, Intricate, Low Detail, dramatic

Sampler: Euler a
Model: epicrealism_newEra



(best quality, masterpiece), 1 boy, traveler, destroyed city, upper body, dark mood, beard, backpack, jumper

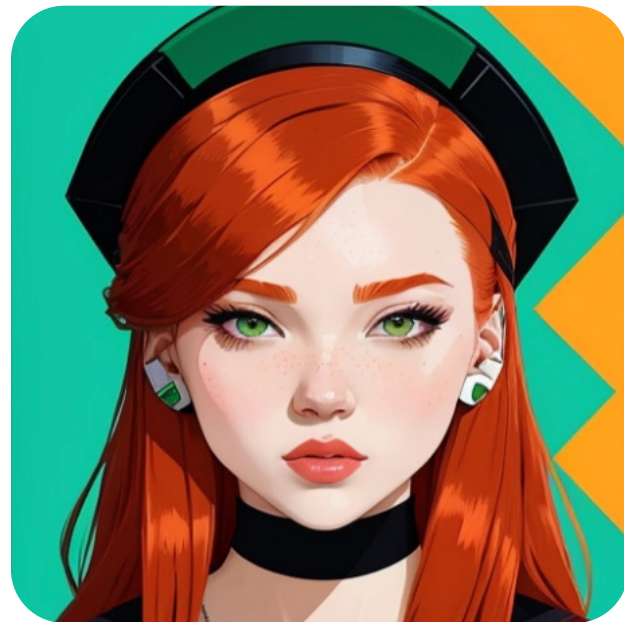
Sampler: DPM++ SDE Karras
Model: toonyou_beta3



3dmm style, portrait of a giant, science fiction, armored, wearing_breastplate, perfect face, pretty face, purple eyes, purple hair, very short hair, flat chest, lush detail, absurdres, <lora:3DMM_V7:1>

Sampler: Euler a
Model: revAnimated_v122

Showcase Prompts



abstract art, 1998 european (redhead:1.1)
hiphop girl (by sachin teng x supreme:1.1),
attractive, stylish, designer, green,
asymmetrical, geometric shapes, graffiti,
street art, urban art

Sampler: DPM++ SDE Karras
Model: dreamshaper_6BakedVae



a professional photography of a skeleton
meditation, deep meditation, sitting cross-
legged, ethereal, hdr, extremely detailed

Sampler: DPM++ 2M Karras
Model: reliberate_v10



an ancient egyptian ritual,, Vector art, Vivid
colors, Clean lines, Sharp edges, Minimalist,
Precise geometry, Simplistic, Smooth curves,
Bold outlines, Crisp shapes, Flat colors,
Illustration art piece, High contrast shadows,
Technical illustration, Graphic design, Vector
graphics, High contrast, Precision artwork,
Linear compositions, Scalable artwork, Digital
art

Sampler: DDIM
Model: colorful surrealismai_V10



a ultradetailed beautiful panting of a
woman dribbling a basketball, by conrad
roset, greg rutkowski and makoto shinkai,
trending on artstation

Sampler: DDIM
Model: deliberate_v2



high quality 3 d render hyperrealist very
cute multipastel dotted fluffy! tarantula
cat hybrid with detailed fluffy wings! !
, vray smooth, in the style of detective
pikachu, hannah yata charlie immer,
dramatic blue light, low angle, uhd 8 k,
sharp focus

Model: deliberate_v2
Sampler: DDIM



A psychedelic portrait of a woman, vibrant
color scheme, highly detailed, in the style
of romanticism, cinematic, artstation,
Moebius, Greg rutkowski

Sampler: DDIM Model: deliberate_v2

Showcase Inspiration

Model: Deliberate V2

Many of the prompts I use originate from the inspiring communities at Civit, HuggingFace, Lexica, and GitHub. While I do customize these prompts to fit specific models, the original credit is due to the incredibly inventive Synthetic Data (SD) community. Their creativity and innovation significantly contribute to the work we do.



Showcase Inspiration

Darth Vader



Showcase Inspiration

Checkpoint: **Colorful Surrealism**



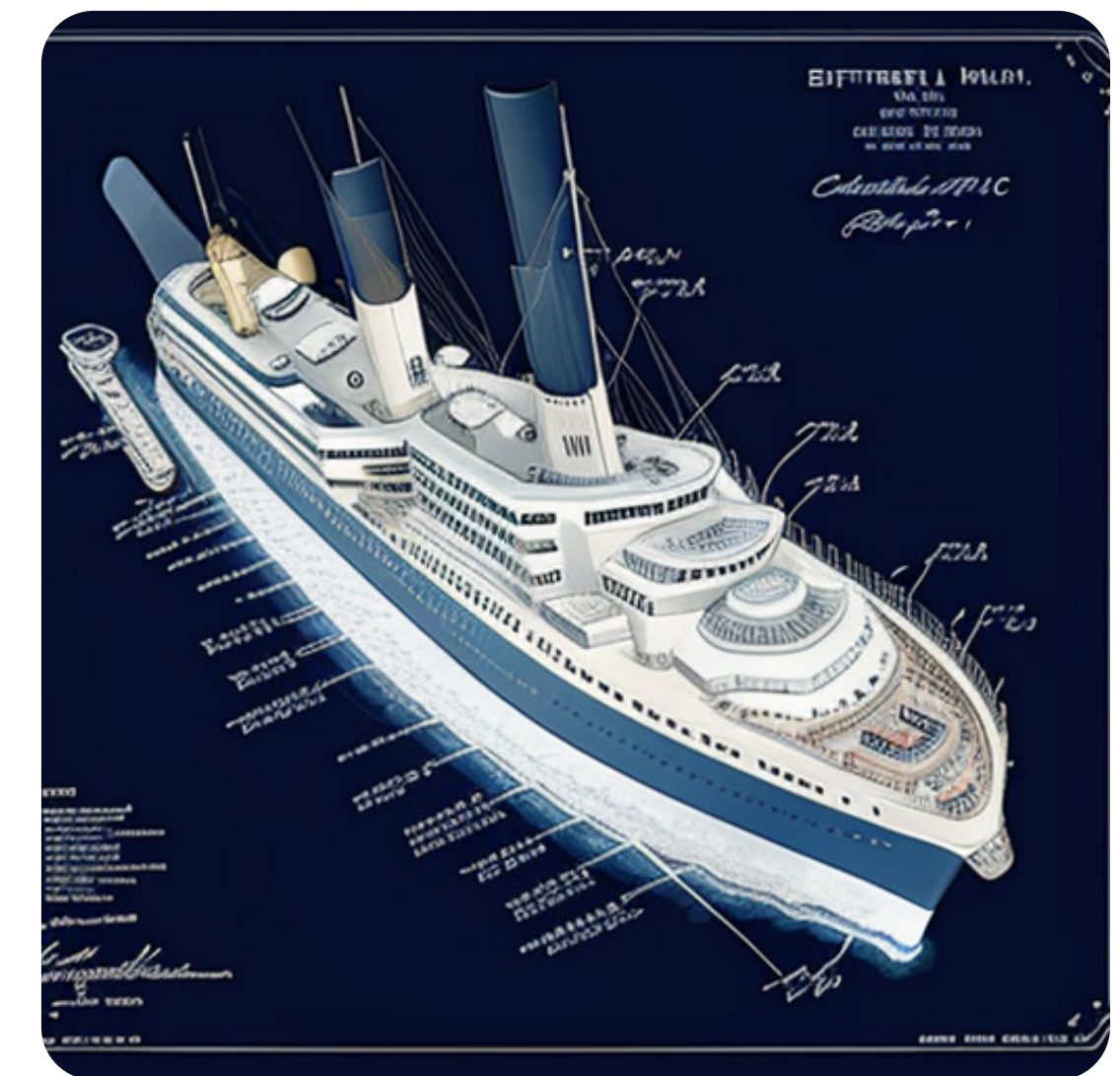
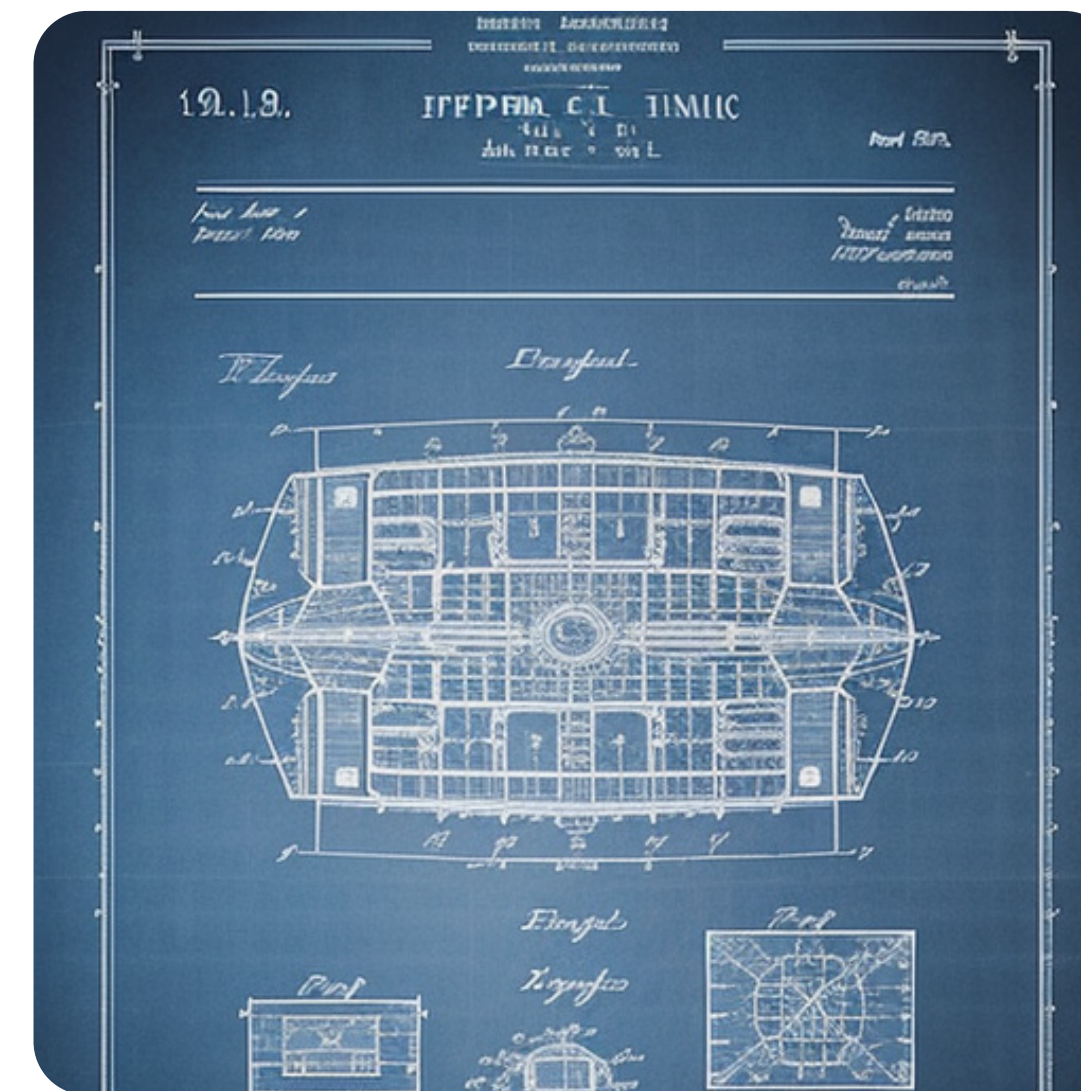
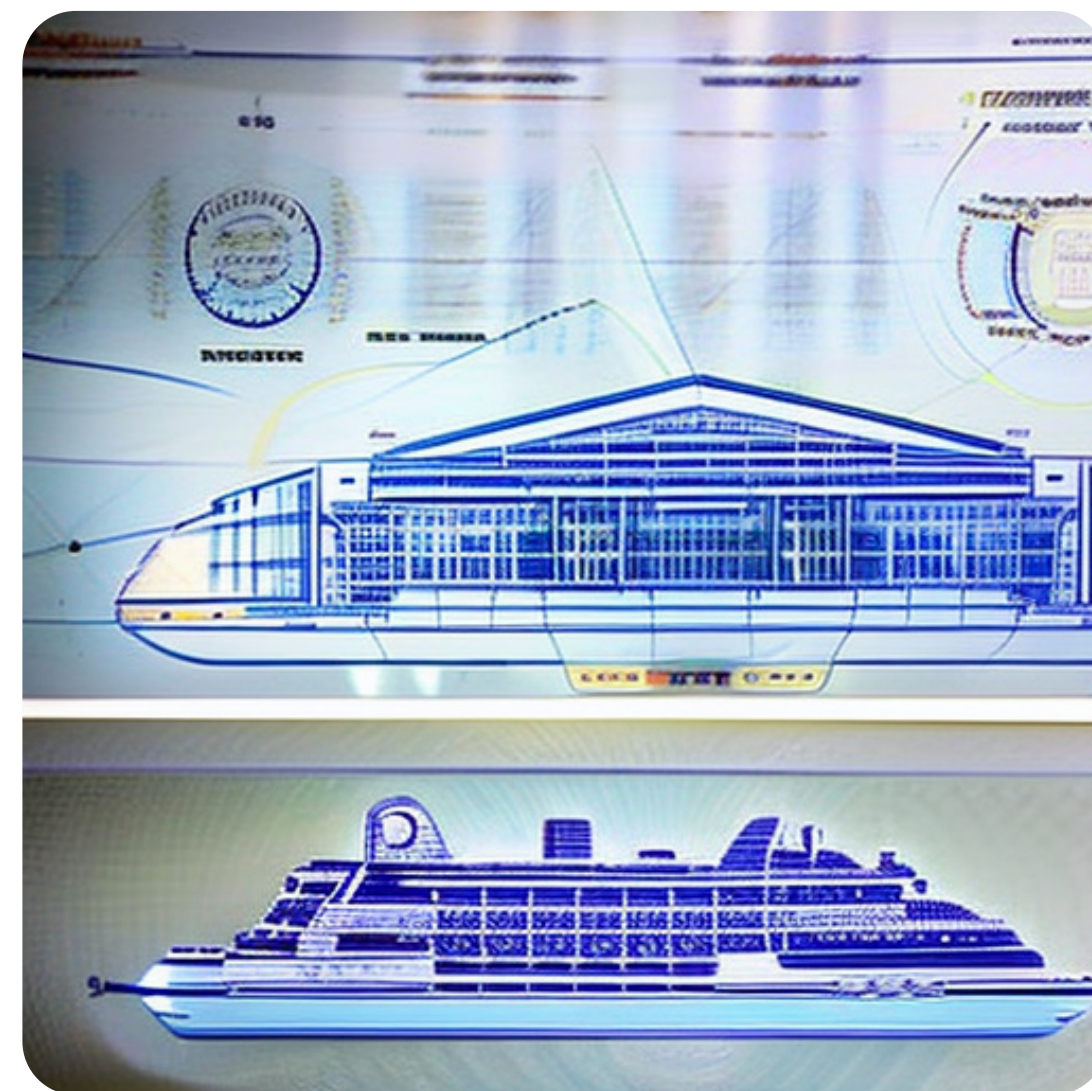
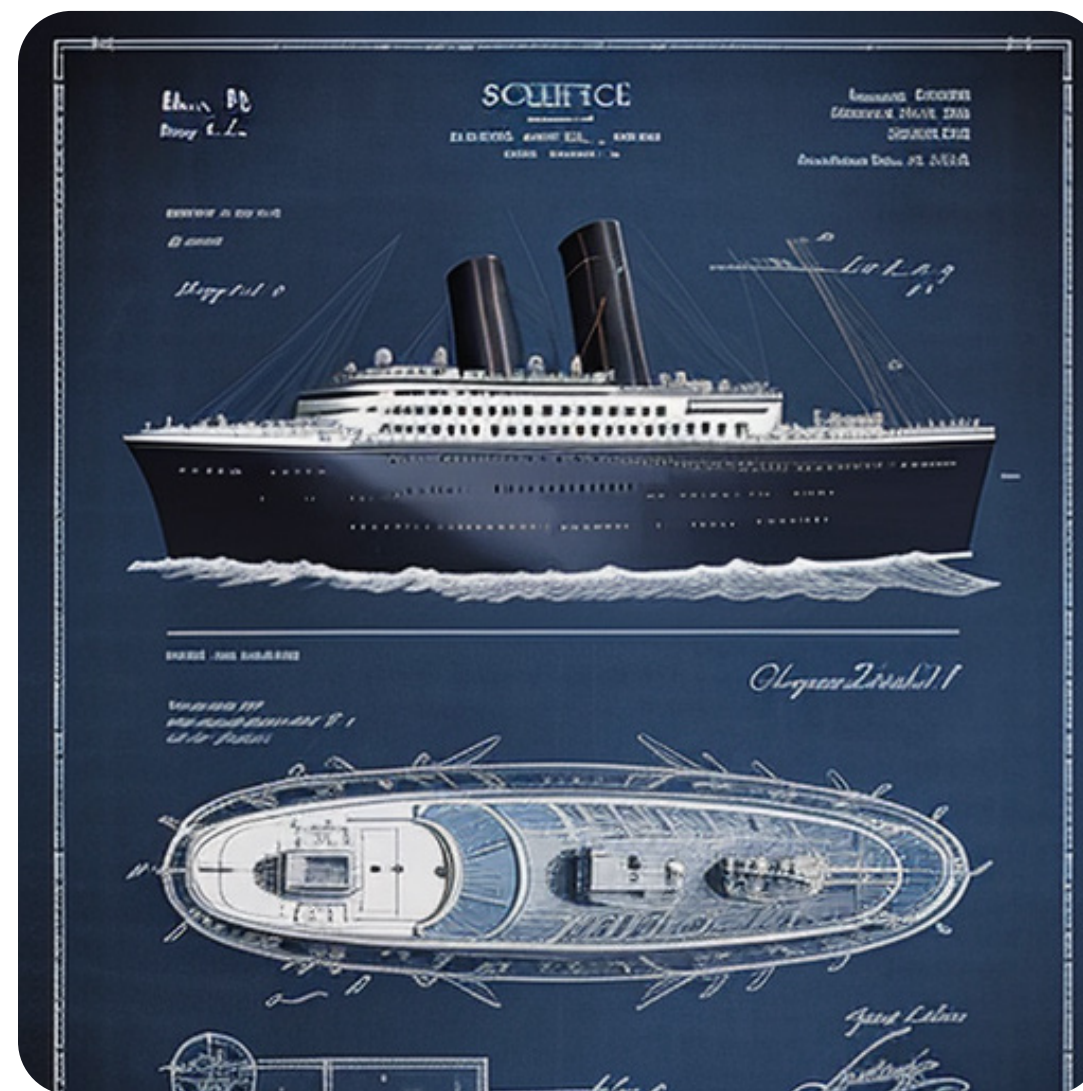
Showcase Inspiration

Foggy Cityscapes



Showcase Inspiration

Blueprints



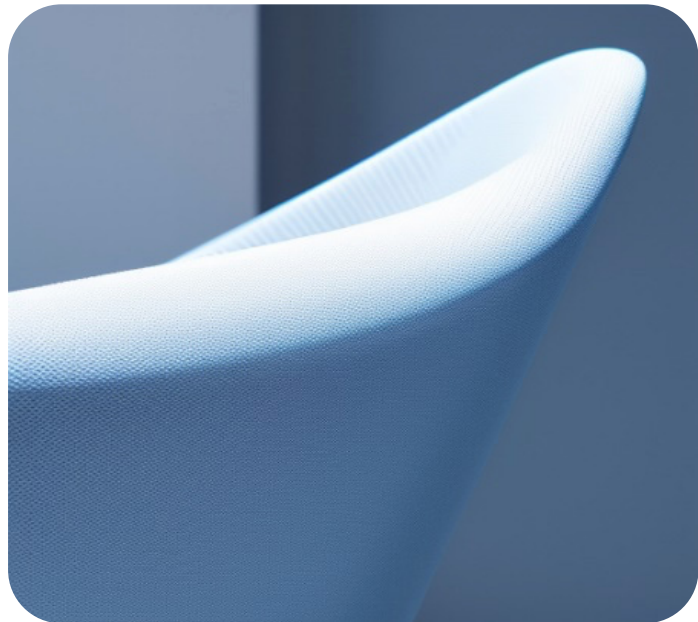
Showcase Inspiration

Product Renders

Product Design:
Lora 1
Lora 2

Product Design:
Checkpoint

Szechuan Special Sauce:
Checkpoint



Showcase Inspiration

Cars



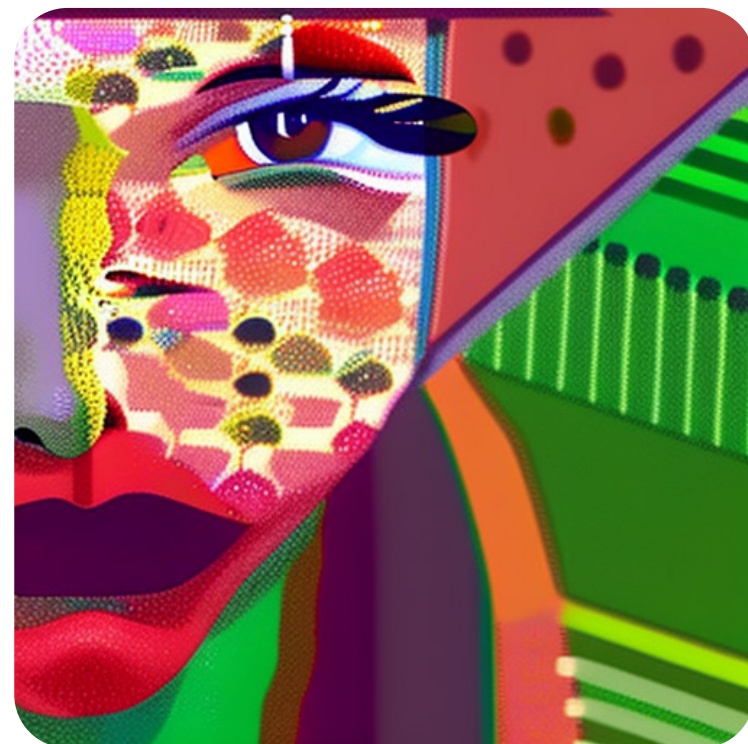
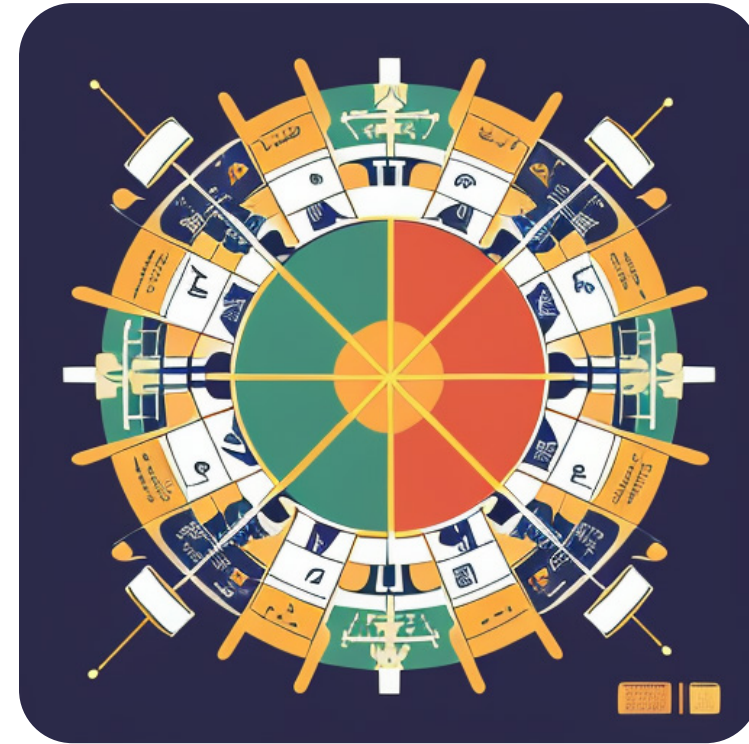
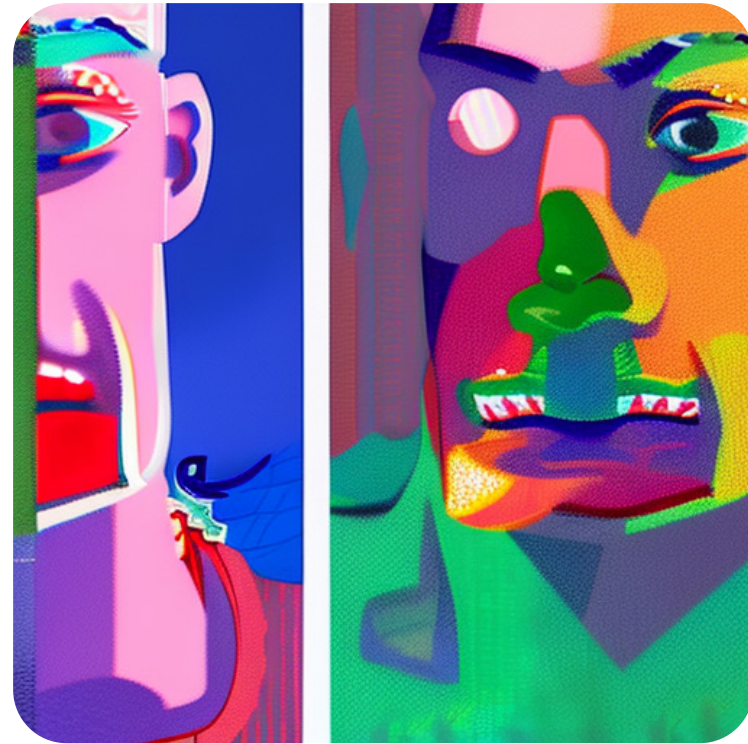
Showcase Inspiration

Cyber Punk



Showcase Inspiration

Abstract Art



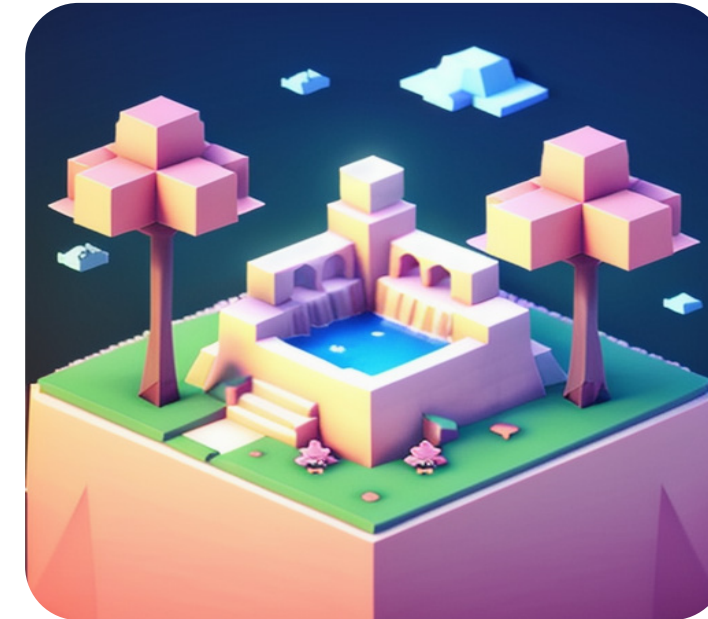
Showcase Inspiration

3D Cabins



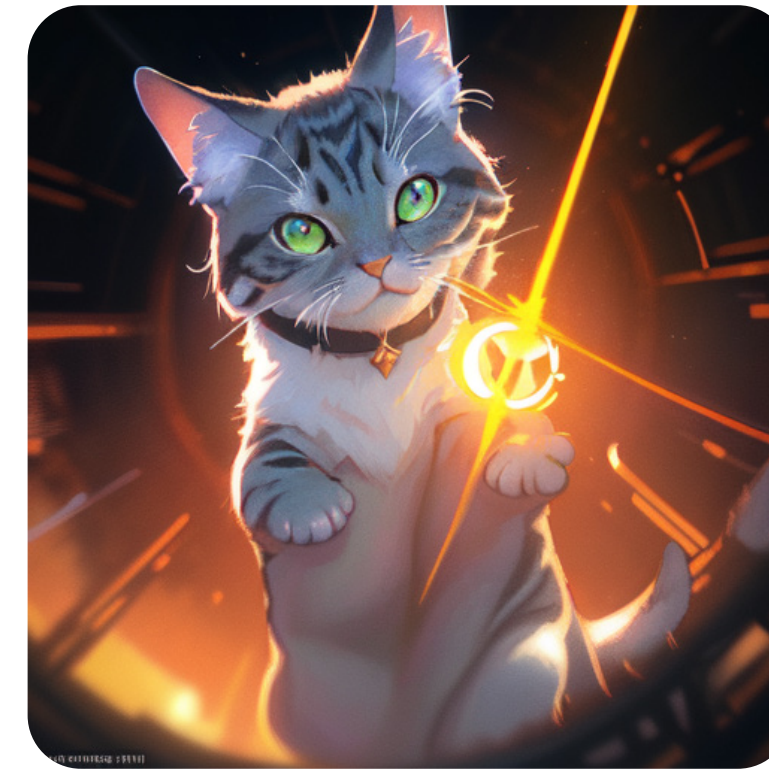
Showcase Inspiration

Isometric



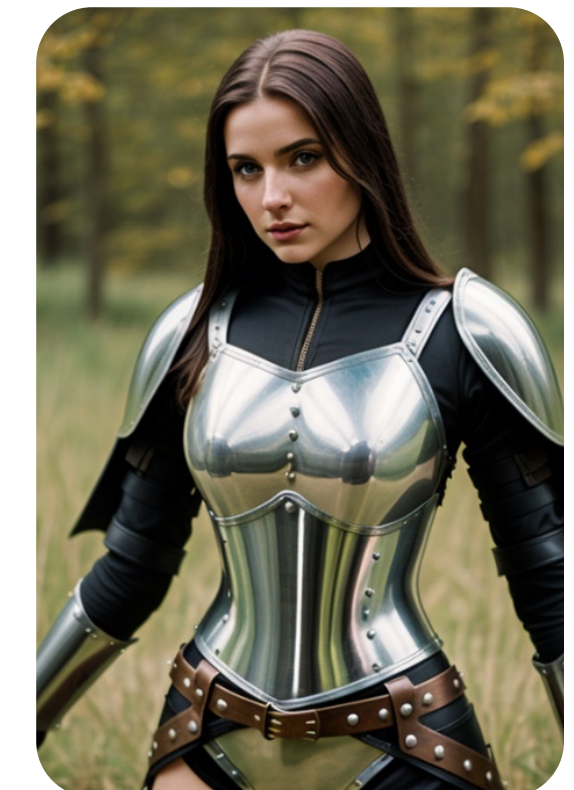
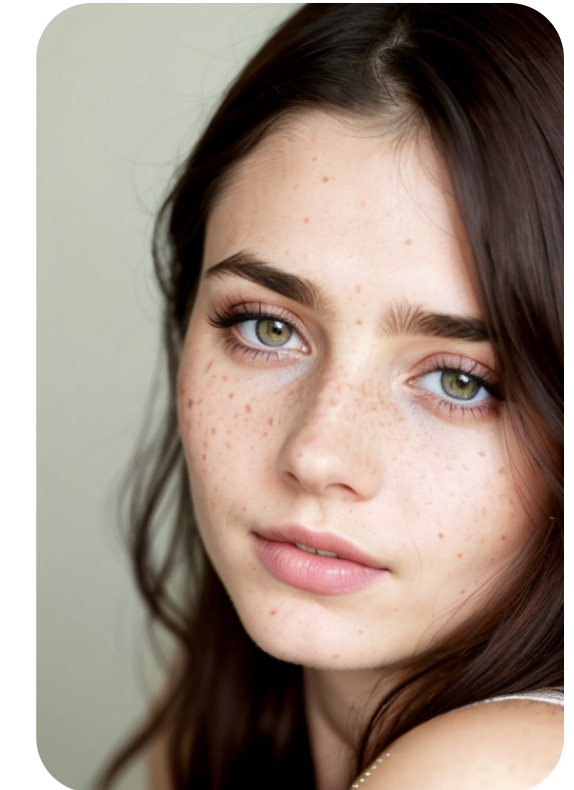
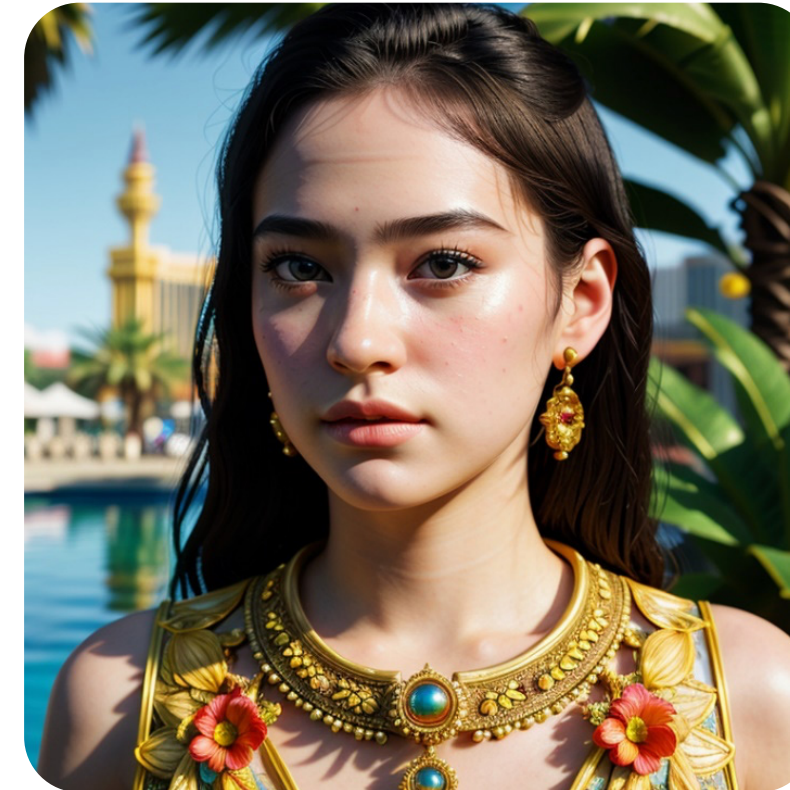
Showcase Inspiration

Cats



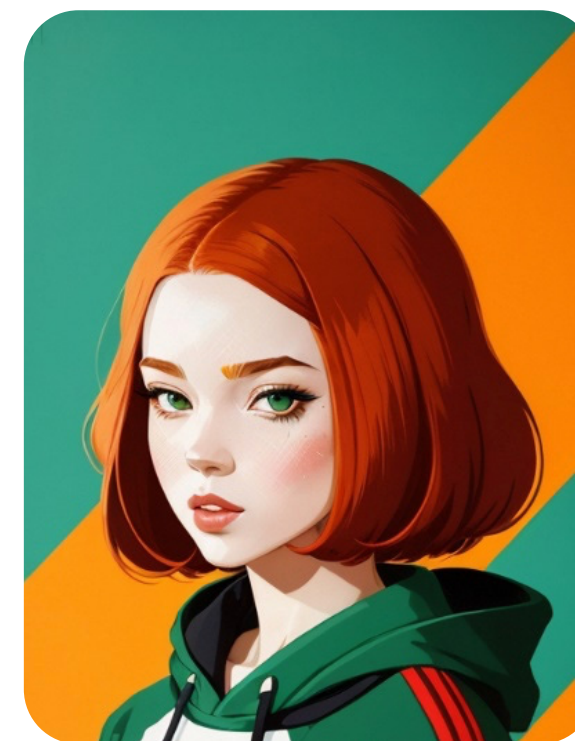
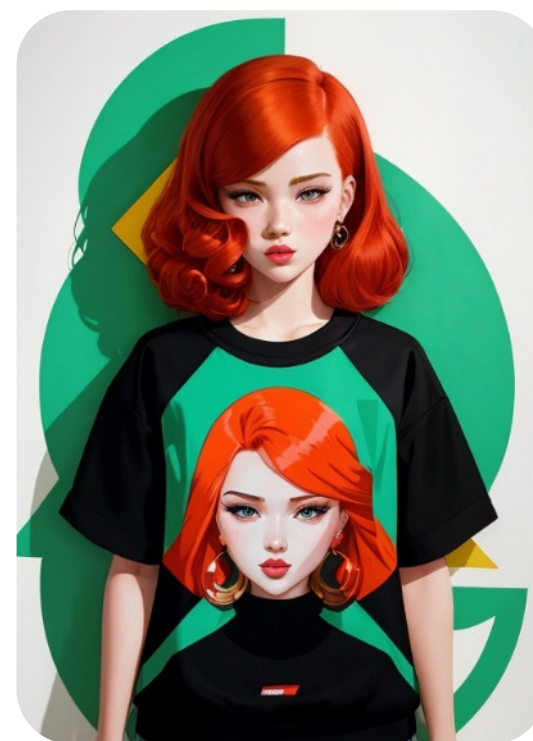
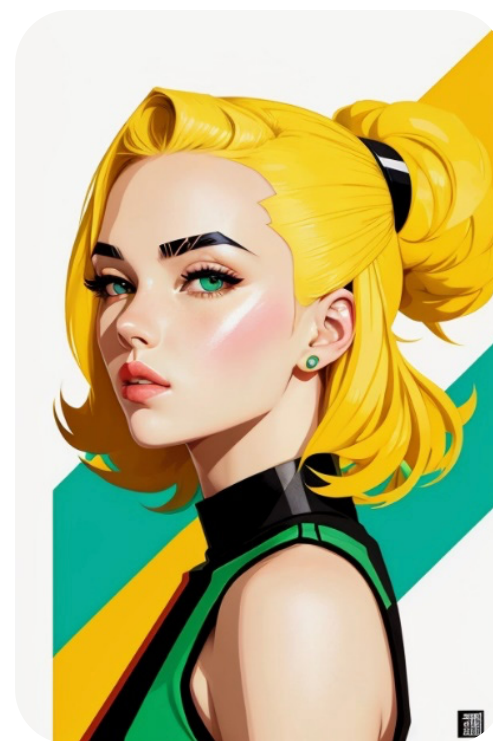
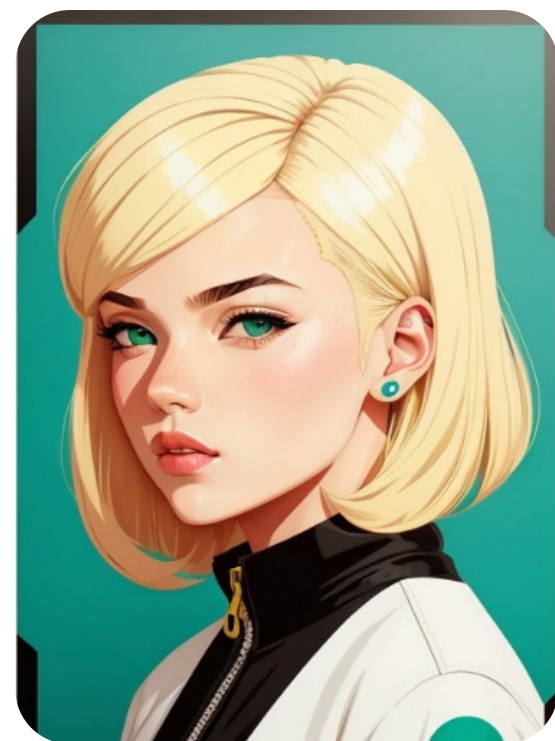
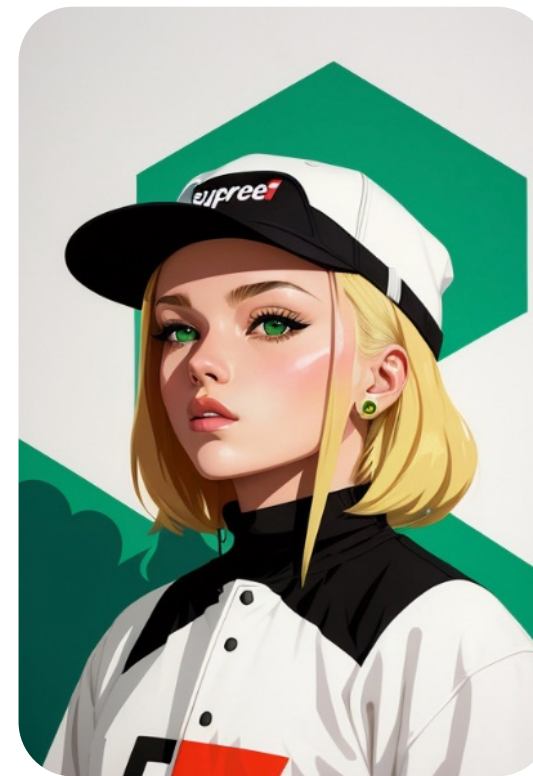
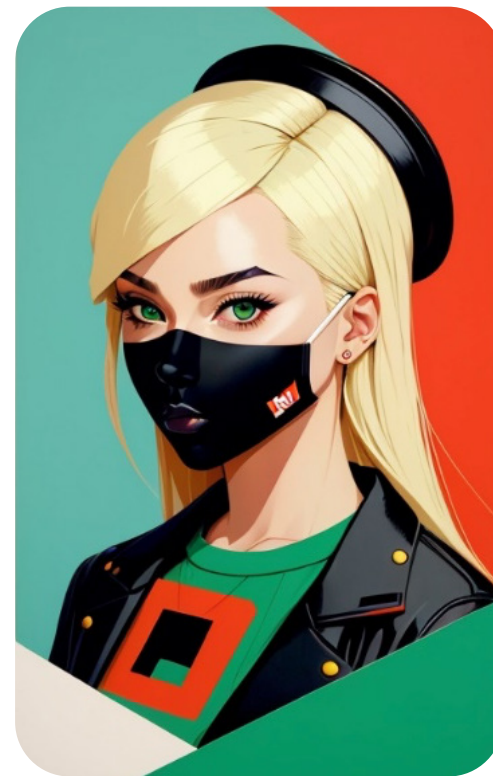
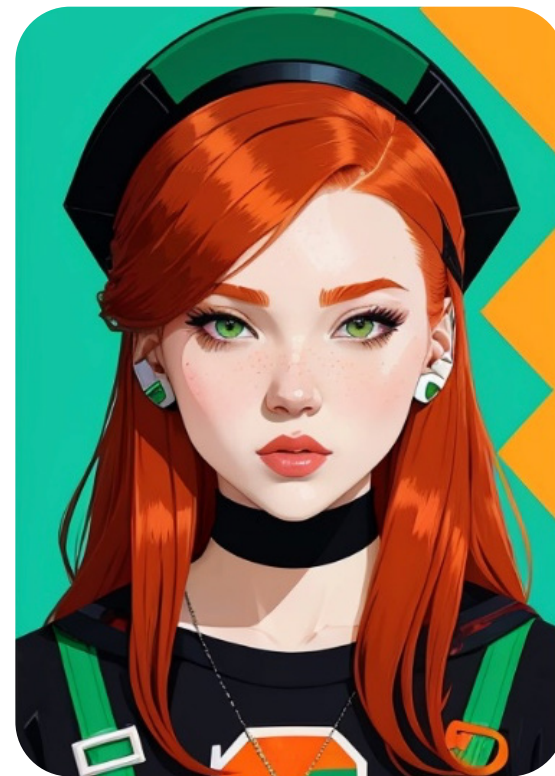
Showcase Inspiration

Realistic



Showcase Inspiration

Style-specific Characters



Showcase Inspiration

Food



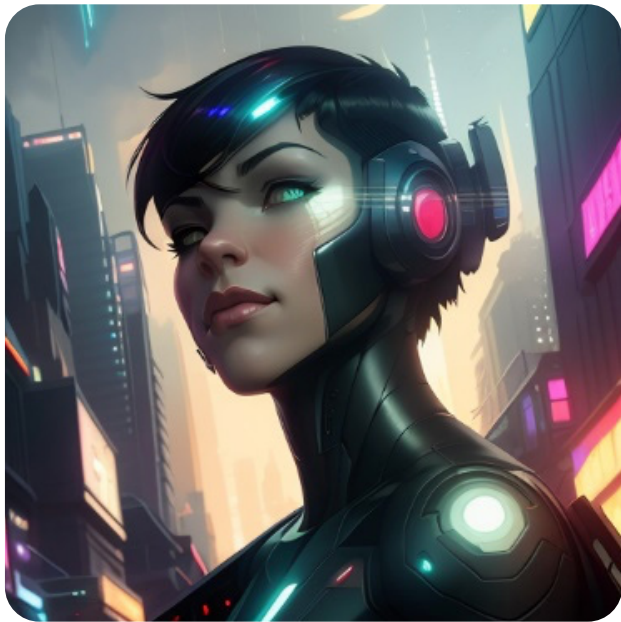
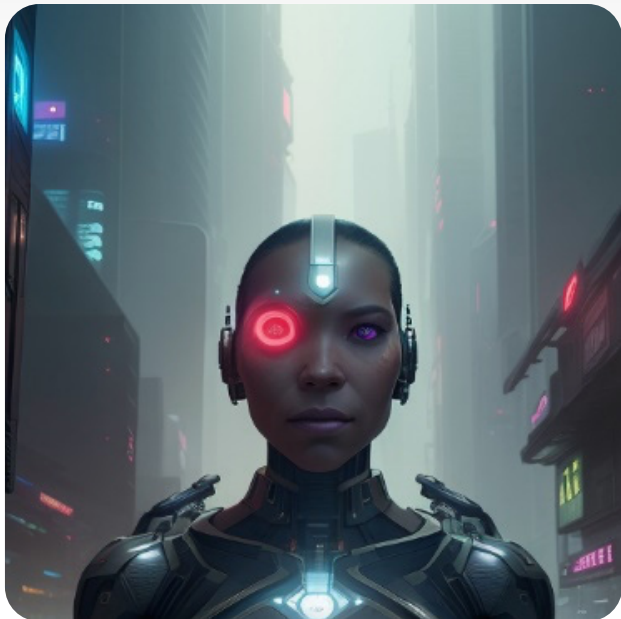
Showcase Inspiration

Landscape Illustration



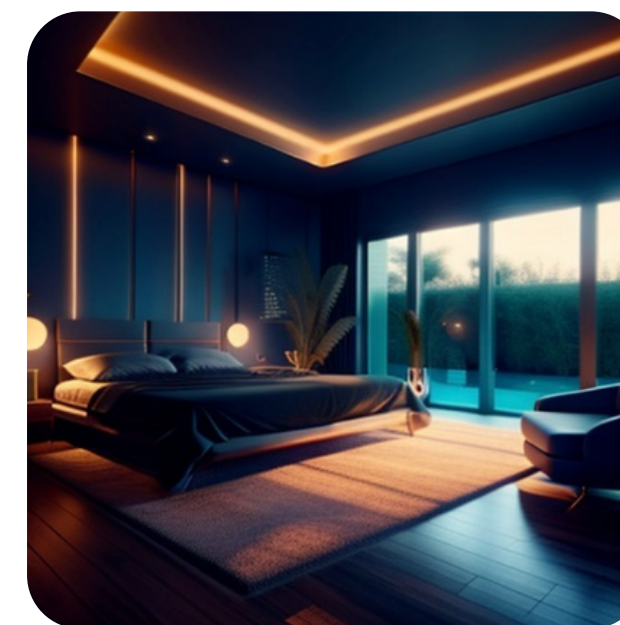
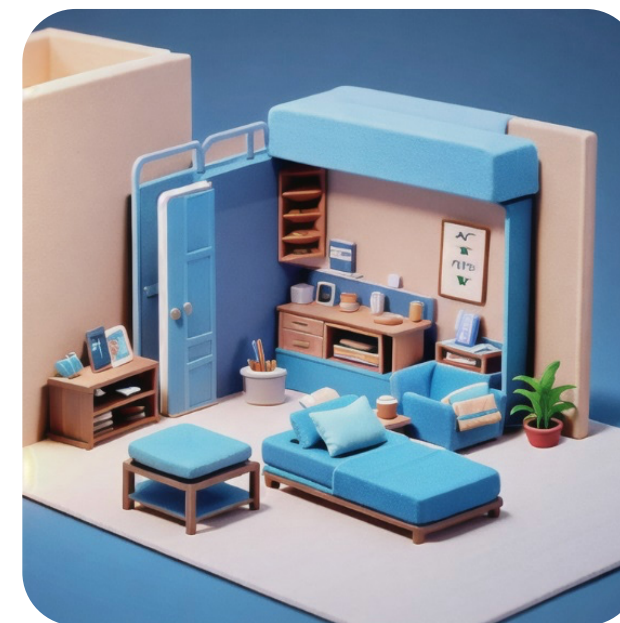
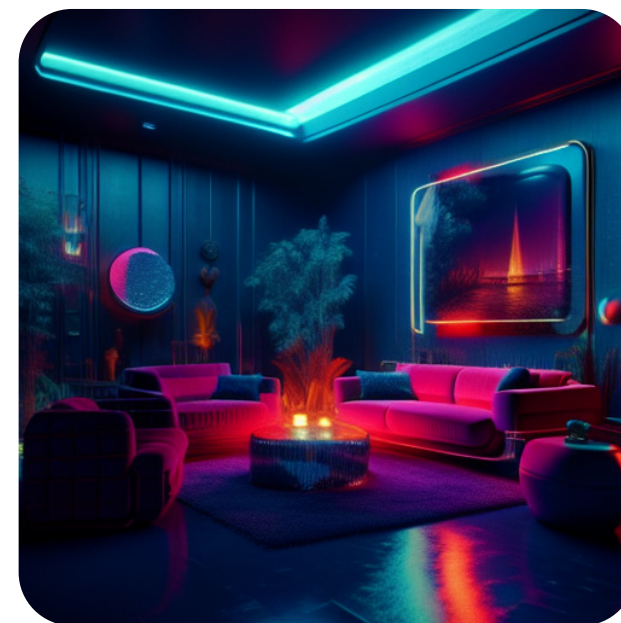
Showcase Inspiration

Dystopia



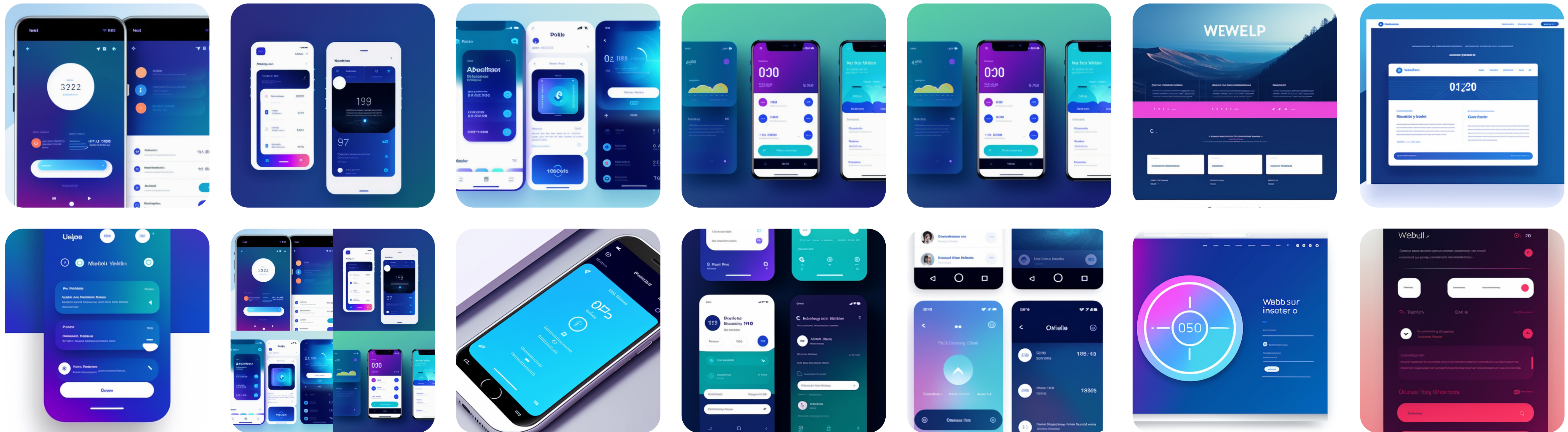
Showcase Inspiration

Interior design



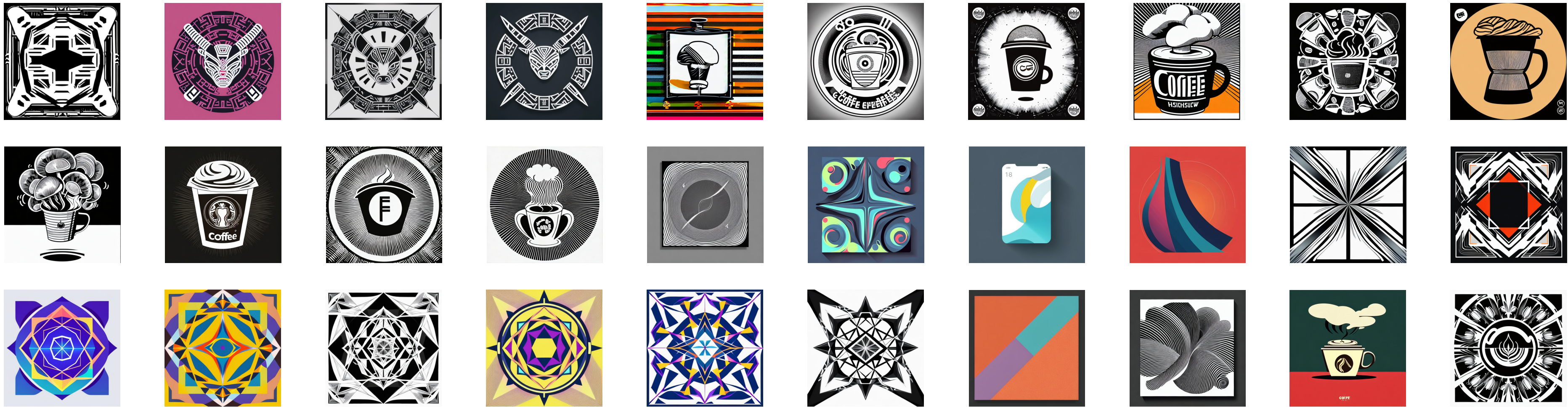
Showcase Inspiration

Web UI



Showcase Inspiration

Logos



06

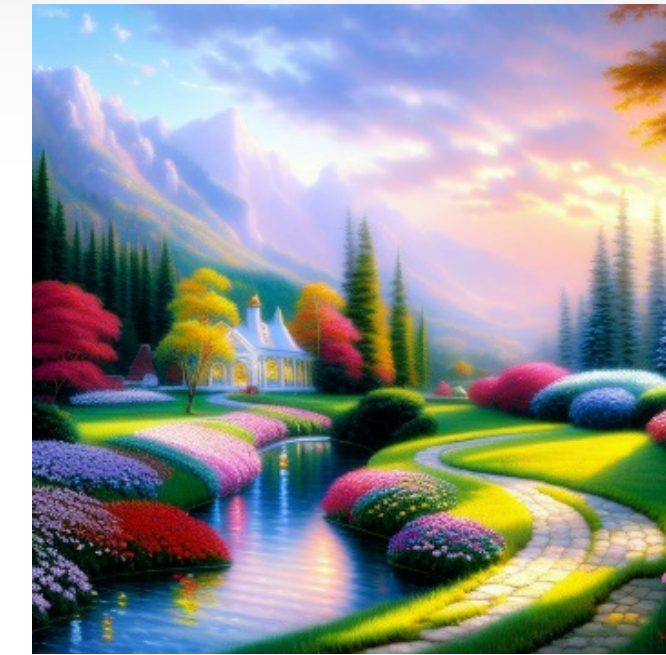
ETHICAL AND LEGAL IMPLICATIONS

Ethical and Legal Concerns

Emerging AI technology for image generation, such as Stable Diffusion, raises both ethical and legal concerns related to copyright and intellectual property rights. By training AI models on copyrighted material, it has the potential to generate images that encroach on existing rights, triggering complex questions on ownership, usage, and legal implications. To navigate this intricate landscape, it's crucial to identify areas of safe and fair use, where generated images avoid referencing real people, identifiable objects, or distinct brands.

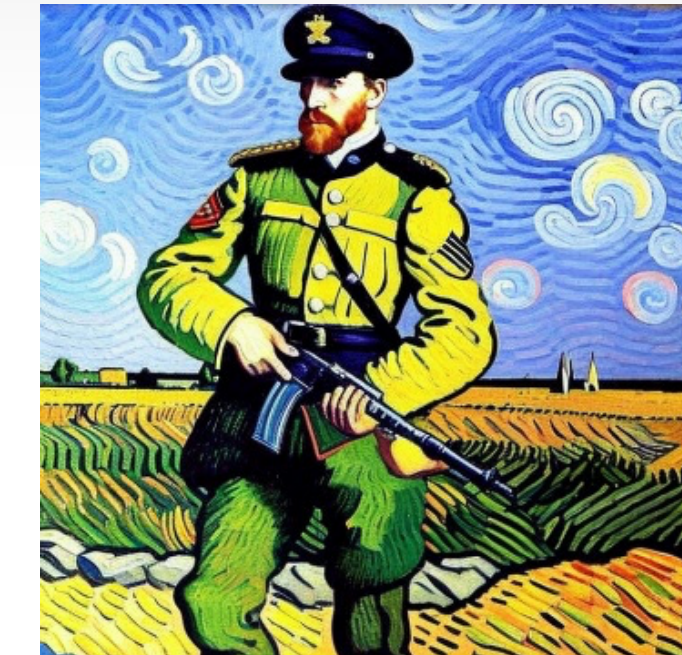
One common use of AI image generation is through platforms like Dreamstudio.ai, where images generated are classified as "public domain," permitting anyone to use these images for any purpose. However, when Stable Diffusion is employed locally on your personal computer, you hold the copyright over the created images, similar to creating an image in Photoshop.

A dilemma arises when generated images are created using prompts that reference known artists or specific styles, for example, "...painting by Rutkowski." The unresolved question is whether Rutkowski can claim any rights over the resulting art. If the objective is absolute safety, it is advised to create images without borrowing the style of any living artist.



Thomas Kinkadee

Prompt: A painting of a heavenly utopia,
by Thomas Kinkadee



Vincent Van Gogh

Prompt: a painting of a modern soldier
by Vincent Van Gogh



Leonid Afremov

Prompt: a painting of Donald duck by
Leonid Afremov



Claude Monet

Prompt: a portrait, of a painting of
Donald trump, by Claude Monet

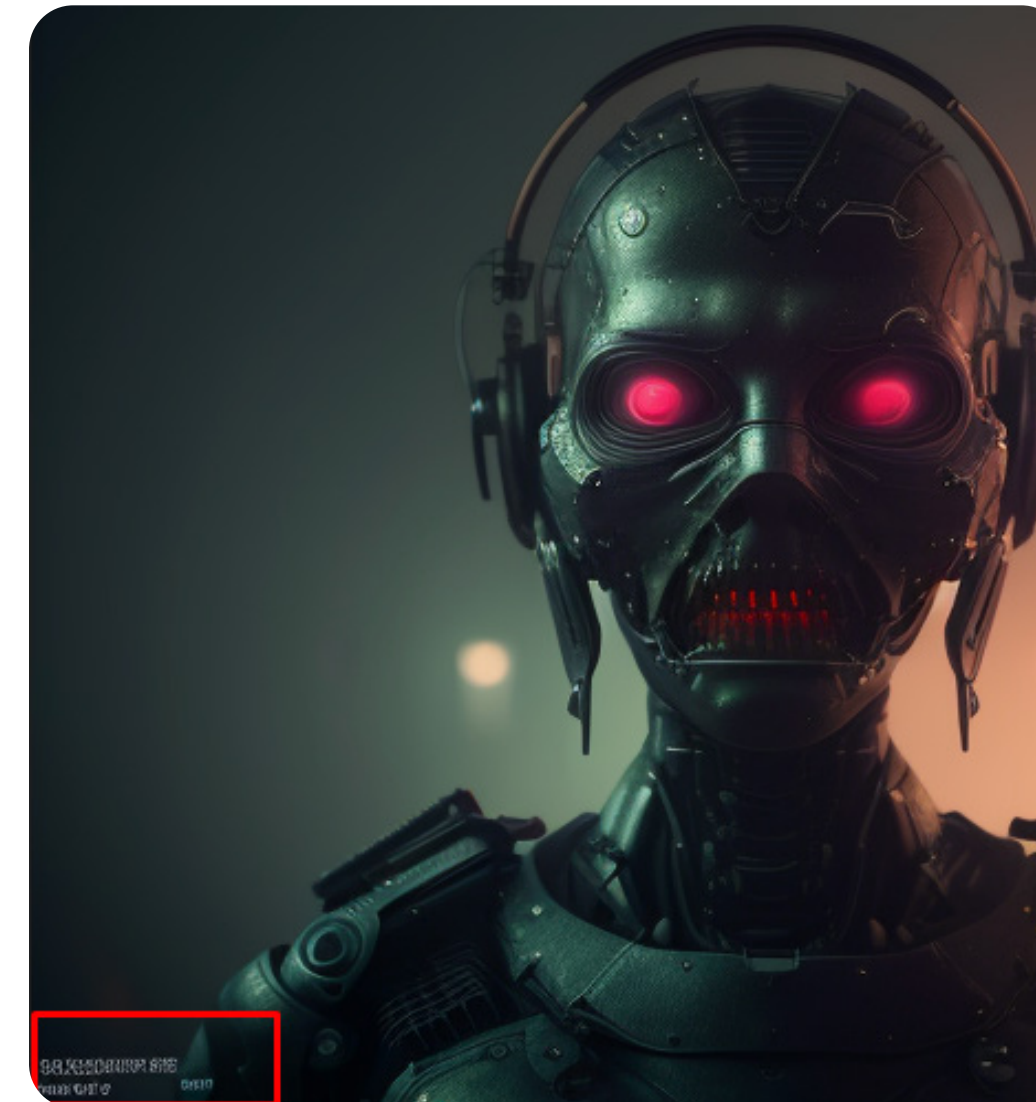
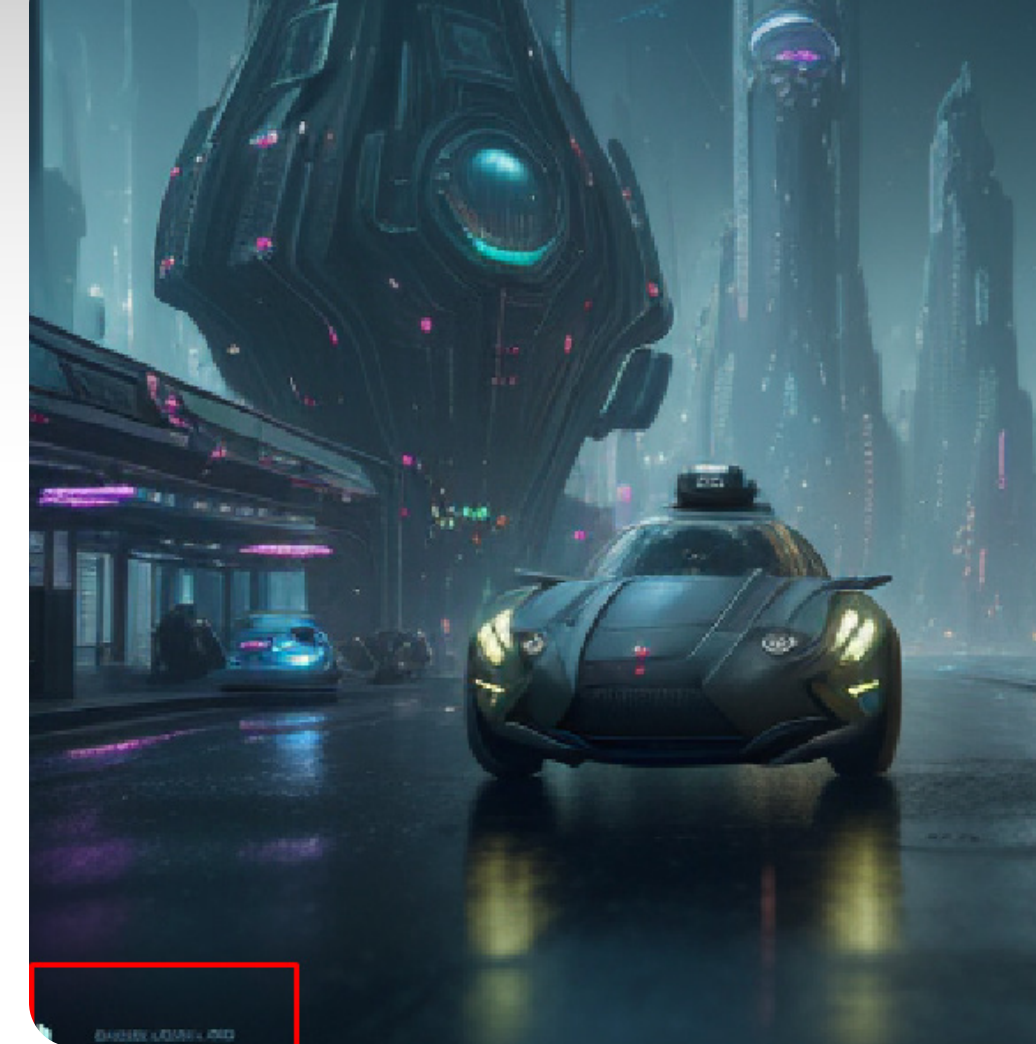
Ethical and Legal Concerns

There are instances where AI-generated artworks bear traces of watermarks and signatures, indicating an imperfect attempt by the AI to either erase or mimic these marks, thus circumventing intellectual property rights.

The U.S. Copyright Office recently set a precedent, refusing to grant copyright for an image created by an AI program, stating that “human authorship is a prerequisite to copyright protection.” The dispute is ongoing, with the AI program’s owner filing an appeal.

Cases like these, including a lawsuit against Stable Diffusion and Midjourney, further demonstrate the unsettled legal landscape around AI image generation. If the court rules that each generated image is partially copyrighted by every individual whose image was included in the training set, it would render the images commercially unviable.

While you are permitted to sell the works you generate, it doesn't grant you the right to sell generated images of copyrighted characters or individuals. Furthermore, you cannot copyright what is produced, meaning others can freely use your generated images.



Ethical and Legal Concerns

The likeness of AI-generated images to images in its training dataset raises significant concerns about copyright infringement. In addition, the generated images may not be copyrightable in your jurisdiction, but that does not bar you from selling them.

On a case-by-case basis, it may be argued that AI-generated images bear substantial similarities to other copyrighted works, potentially leading to drawn-out legal disputes. Furthermore, the inability to protect your images means anyone can copy and sell them, without much legal recourse.

The legal conundrum extends to third-party distributors, who could face liability for distributing potentially infringing content and may encounter difficulties protecting their interests in such works.

Important note:

When using checkpoints, LORAS, or Embeddings, you must always double check to see what their license states, which is usually available on the source website. You're mostly safe to use this content under "fair-use" but it's always best to double check, especially if you intend to use the generated images for your marketing material.



Ethical and Legal Concerns

Using AI to generate images of celebrity lookalikes or mimic distinct designs could potentially expose you to legal action, particularly from individuals and companies known for their litigiousness. This is why it's essential to exercise caution and common sense, particularly when dealing with high-value images or dealing with highly-emotional subjects.

Understanding the concept of "fair use," a legal doctrine allowing the use of copyrighted material without permission under certain limited circumstances such as criticism, comment, news reporting, teaching, and research, can help you navigate the legal intricacies surrounding AI image generation.

In our current reality, machines are fed millions of pieces of human creativity, generating new works from this cumulative data. This development has led to an ethical and legal gray area, with stakeholders across industries offering diverse perspectives on the implications for their work and the nature of art itself.

To maintain ethical and legal safety, you could consider using base images that you own, generate singular generic elements, or use AI-generated images as inspiration for your design process. Keeping abreast of evolving laws and cases around the world can also help you understand how these concerns will be addressed over time.

However, it is advisable to steer clear of certain actions. These include reselling, posting, or marketing content that closely resembles another artist's work or references a particular style such as Disney or Pixar, which should be restricted to personal entertainment or inspiration. Images that bear a striking resemblance to human figures, particularly public figures, should also be avoided.

Exploitation of this technology for illegal activities or content creation, such as deepfakes or NSFW content, can lead to harmful or unethical outcomes and should be strictly avoided. Furthermore, using any likeness of public figures for commercial gain without their consent can potentially lead to litigation, as there are instances where celebrities and companies have aggressively defended their image rights.

Ultimately, while AI image generation models like Stable Diffusion open a new world of possibilities, they also usher in complex ethical and legal concerns. As we continue to explore this exciting frontier, it is critical to tread carefully, respecting existing intellectual property rights and norms, and to be mindful of the potential for misuse.

The coming years will likely bring new laws, precedents, and norms to this rapidly evolving field. Until then, prudence, respect for others' work, and a keen eye on emerging legal developments are key to harnessing the power of AI image generation responsibly.

Indeed, AI's potential is as potent as the wisdom with which we wield it. With a bit of common sense and cautious respect for copyright and ethical guidelines, we can ensure this technology is used in a manner that both respects the rights of individuals and promotes continued innovation in the field.

Installation

General information and benefits of stable diffusion online and on your device include:

Stable Diffusion operates on an open-source basis, making its underlying code accessible for free. This allows users the chance to not only access and alter the code, but also contribute to its development, fostering an atmosphere of innovation and collaboration within the AI community.

Users of Stable Diffusion are provided with rights to the images they generate. This ensures that users have full ownership and control over the content they create, thereby providing them with the flexibility and freedom to use their images in any way they choose.

Users can access Stable Diffusion online through a variety of platforms. Additionally, certain websites offer free access to Stable Diffusion. For example, rundiffusion.com is recommended for those seeking a simplified version of Stable Diffusion, while playgroundai.com is even more user-friendly and is thus recommended for beginners.

A number of online platforms utilize Stable Diffusion, including Mage, Hotpot.ai, StableDiffusionOnline.net, and StableDiffusionWeb.com. These platforms offer a range of benefits, from easy and fast access to Stable Diffusion for generating a wide range of images (Mage), to enhanced versions of Stable Diffusion that are optimized for user-friendliness and speed (Hotpot.ai).

Installing Stable Diffusion directly onto your device yields additional advantages. By doing so, you can gain full control over the model's parameters, allowing for customization and refinement based on your specific needs. This degree of control permits users to tailor the generated images to their preferences and the requirements of their projects.

Furthermore, installing Stable Diffusion on your device allows you to generate a larger quantity of images without any restrictions typically imposed by online platforms or services. This means you can truly explore the full capabilities of Stable Diffusion and generate a substantial amount of content for a variety of applications.

Lastly, by installing Stable Diffusion on your device, you can alleviate concerns regarding censorship or access restrictions that may arise when utilizing online platforms. With the model installed directly on your device, you can bypass any potential limitations and ensure consistent use.

In conclusion, Stable Diffusion has numerous advantages. It operates on an open-source basis, allows users to retain rights to the images they generate, and provides a variety of options for online use. Additionally, installing Stable Diffusion directly on your device allows for full control over its parameters, the ability to produce a higher number of images, and the avoidance of potential censorship issues.

Hardware Specifications Required:

For an optimal experience with Stable Diffusion, your device should have a quality GPU and a separate GPU. You can verify your GPU capabilities by pressing Control + Shift + ESC to open the Task Manager, then navigate to the Performance tab and check your GPU status. The minimum requirement is 4GB of dedicated GPU memory. Also, ensure that you have at least 10GB of available storage on your hard drive.

Important Updates:

Stable Diffusion Version 2.0:

The open-source version of the model, Stable Diffusion v2.0, marked a significant milestone. This release greatly influenced the open-source AI model community, sparking the creation of numerous models and groundbreaking developments.

Stable Diffusion Version 2.1:

Following the release of Stable Diffusion v2.0, version 2.1 was introduced. The aim of v2.1 was to resolve some of the v2.0's relative weaknesses, especially those related to the modified NSFW (Not Safe for Work) filter.

Installation

Let's begin with the required installations - Python and GIT:

However, a word of caution. As of May 2023, there are some online resources which, although still functioning, might be a bit outdated, potentially leading to some confusion. If you're interested in video tutorials, I recommend checking out recently posted videos by well-known YouTubers.

A case in point is this source:

<https://semicolon.dev/stablediffusion/local-install-windows>

One of my colleagues had a hard time navigating it before approaching me for help.

Now, let's start the process:

You'll need to install Python 3.10.6 - this is the version we recommend for this application. Keep in mind, Stable Diffusion is primarily written in Python, a versatile programming language extensively utilized in various domains such as AI, Natural Language Processing (NLP), Machine Learning (ML), and more.

Easy Mode for Windows:

Quick Install Guide: A1111 WebUI Easy Installer and Launcher

If you're looking for a simplified installation process, consider using the "A1111 WebUI Easy Installer and Launcher" from Mozoloa on GitHub. This tool guides you step-by-step through the installation process, automatically downloading and installing all necessary resources in the correct version.

After a successful installation, the user-friendly launcher allows you to manage additional settings, such as enabling automatic updates. This allows you to easily get started with Stable Diffusion without the need for in-depth setup knowledge.

Please note that this book also provides a comprehensive manual installation guide for those who prefer a more hands-on approach or require greater customization.

[Get Installer on GitHub](#)

Installation

You can download Python 3.10.6 from the following link:

<https://www.python.org/downloads/release/python3106-/>

The page provides several releases for different operating systems. Simply scroll down to the bottom until you spot the options illustrated in the image below.

Files					
Version	Operating System	Description	MD5 Sum	File Size	GPG
Gzipped source tarball	Source release		d76638ca8bf57e44ef0841d2cde557a0	25986768	SIG
XZ compressed source tarball	Source release		afc7e14f7118d10d1ba95ae8e2134bf0	19600672	SIG
macOS 64-bit universal2 installer	macOS	for macOS 10.9 and later	2ce68dc6cb870ed3beea8a20b0de71fc	40826114	SIG
Windows embeddable package (32-bit)	Windows		a62cca7ea561a037e54b4c0d120c2b0a	7608928	SIG
Windows embeddable package (64-bit)	Windows		37303f03e19563fa87722d9df11d0fa0	8585728	SIG
Windows help file	Windows		0aee63c8fb87dc71bf2bcc1f62231389	9329034	SIG
Windows installer (32-bit)	Windows		c4aa2cd7d62304c804e45a51696f2a88	27750096	SIG
Windows installer (64-bit)	Windows	Recommended	8f46453e68ef38e5544a76d84df3994c	28916488	SIG

The page provides several releases for different operating systems. Simply scroll down to the bottom until you spot the options illustrated in the image below.

If you've previously installed any versions of Python on your computer, it would be best to remove them. This can be done by navigating to your Control Panel > Add or Remove Programs > Python > Remove.

To verify that Python has been installed successfully, you can check through the Command Prompt (CMD). Access it by searching for 'CMD' in your Start menu or by typing 'RUN > CMD'.

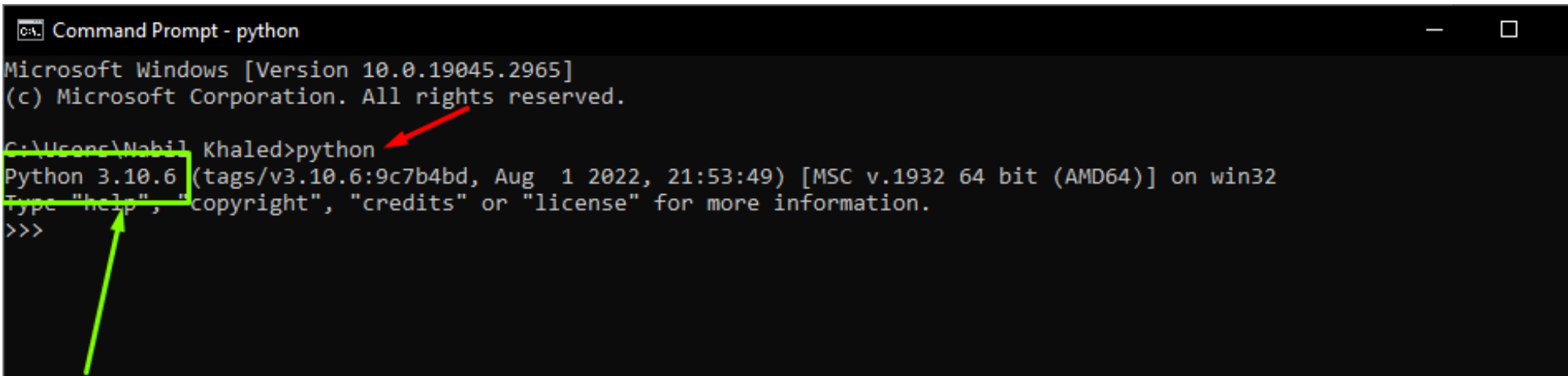
Upon opening the black terminal window, simply type 'python' and press enter. This will display the installed version of Python, confirming that the installation was successful.

Here's a visual guide: In the image attached, I typed 'Python' where the red arrow is pointing, and the green box highlights the output after pressing enter, essentially showing that Python was installed correctly and indicating the version.

Please bear in mind that if this step isn't completed successfully, you need to halt the process. Go back, uninstall Python if it's present on your computer, then reinstall it. Repeat the above steps from the beginning until you're certain the software has been installed correctly.

In case you've installed Python from the Windows Store, I recommend you remove it completely and reinstall it by downloading the installer from the official Python website. In my experience, I had to uninstall and reinstall the software when my first attempt didn't go as planned.

Don't forget to reboot your PC after installing or uninstalling the software. It's a crucial step to ensure the changes are applied properly.



Before proceeding further, it's noteworthy that while some users may prefer alternate command-line interfaces like Microsoft PowerShell, for the purpose of our guide, we suggest using the default Command Prompt (CMD) built into all Windows operating systems.

Installation

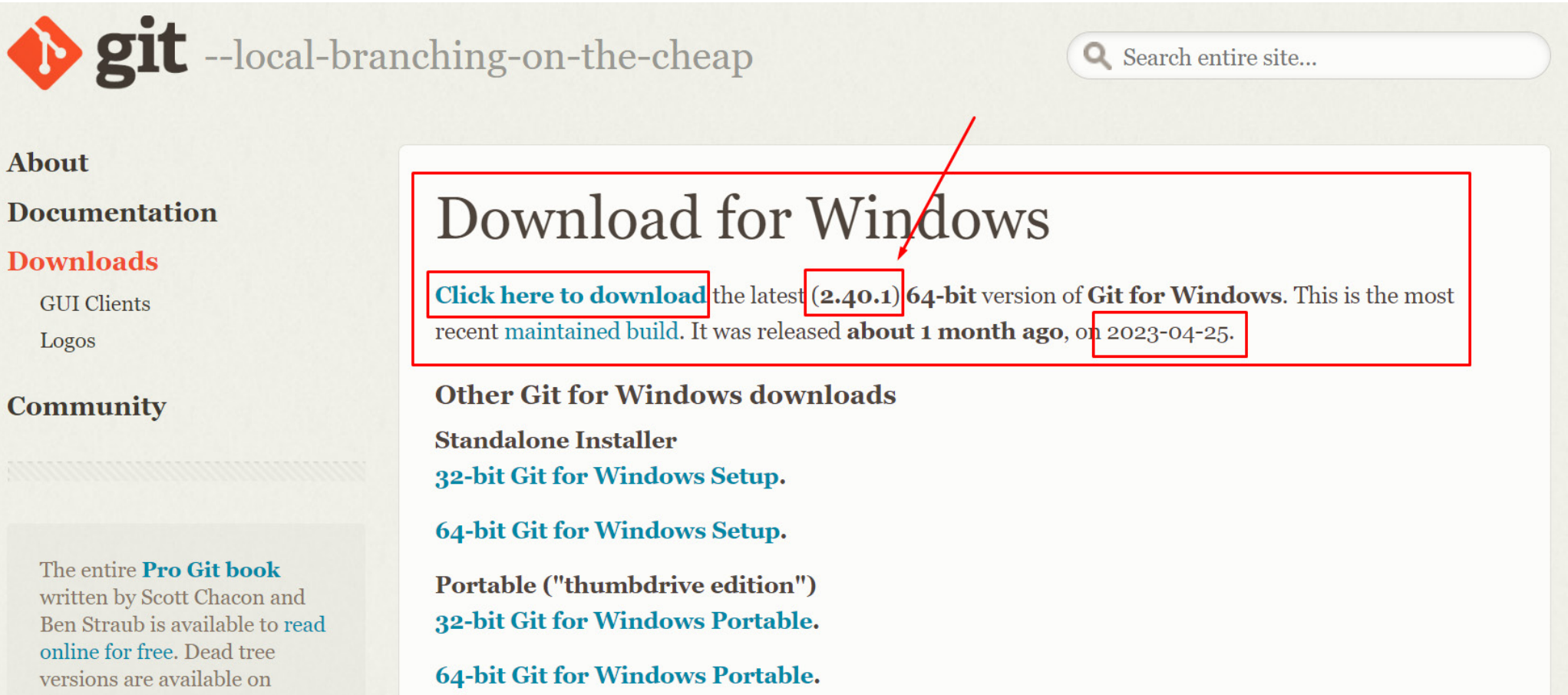
Installing Git

Keep in mind that such information can evolve over time. To ensure the information hasn't become obsolete, it's always a good idea to validate it against popular and trusted sources such as Github, Reddit or regularly cited [websites like https://stable-diffusion-art.com/](https://stable-diffusion-art.com/).

To install Git, navigate to <https://git-scm.com/download/win>.

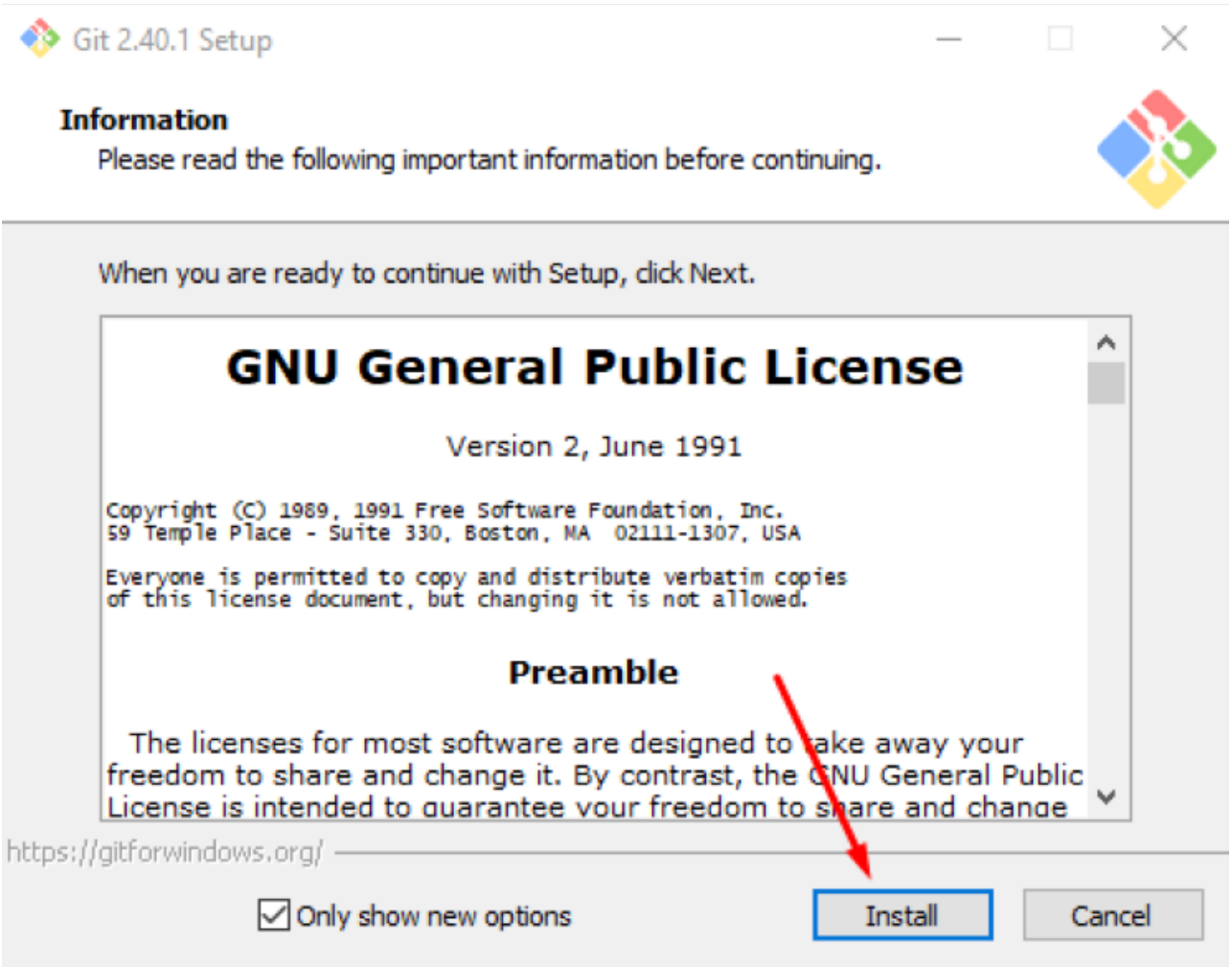
In my case, I bypassed all the options on the page, opting to directly hit the prominent "click to download button" you'll notice at the top left (for version 2.40.1). This works perfectly for me, running a x-64bit system.

Refer to the image below for a better understanding.



GIT is a necessary installation because it enables the set up of AUTOMATIC1111. AUTOMATIC1111 serves as a Graphical User Interface (GUI) for Stable Diffusion - essentially, it's the platform where you'll create your prompts and enjoy all the interesting things you'll learn in this guide. So, it's an essential tool.

Download GIT from <https://git-scm.com/download/win>. This installation process is straightforward - simply keep clicking 'Next' until you reach the end. And don't worry, we've had one of our experts scrutinize the finer details - your data is safe.



Installation

Installing Git

The upcoming step might be straightforward for tech-savvy individuals, but if you're not accustomed to navigating through folders via the Command Prompt (CMD), it might seem a bit challenging. However, by following the next steps attentively, you'll sail through it smoothly.

1. Revisit your CMD (as previously explained) by clicking on your Start menu, typing in CMD, and hitting enter. Once you're greeted with the familiar black screen, copy/paste the following command:

`cd %userprofile%`

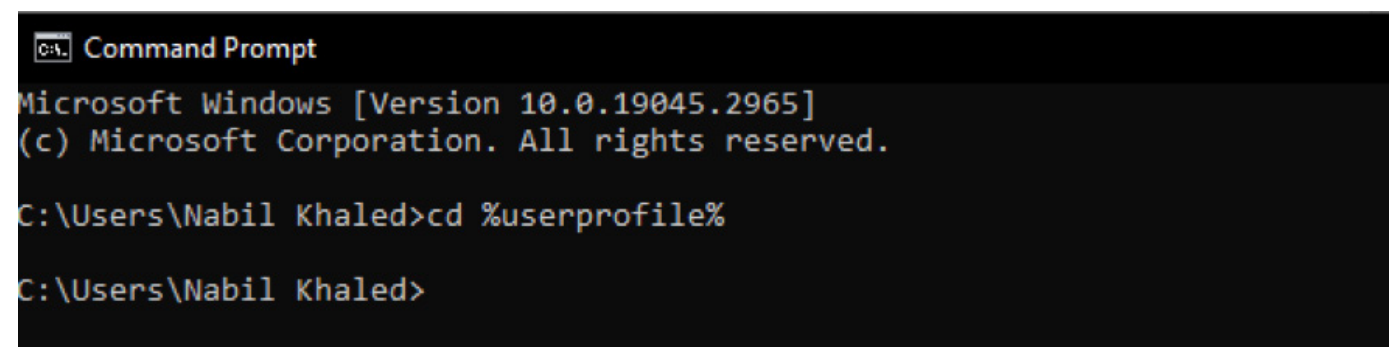
You can right-click and select paste, or use the keyboard shortcuts Ctrl + C to copy and Ctrl + V to paste. If you're reading this guide digitally, you can copy it directly from here.

Upon typing this command and pressing enter, you can be certain that you're in your home folder. The display will look something like this:

`C:\Users\username`

In my case, as you can see in the image below, it shows my username:

`C:\Users\Nabil Khaled>`



```
Command Prompt
Microsoft Windows [Version 10.0.19045.2965]
(c) Microsoft Corporation. All rights reserved.

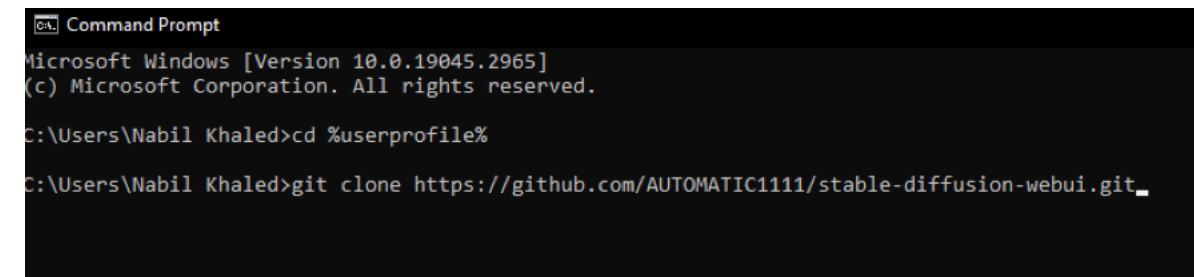
C:\Users\Nabil Khaled>cd %userprofile%

C:\Users\Nabil Khaled>
```

You might want to visit <https://github.com/AUTOMATIC1111/stable-diffusion-webui> to explore more about the features offered by this particular 'fork' (the most popular one currently).

Next, copy and paste the following command, and press enter. This step will clone the AUTOMATIC1111 repository to your local system:

`git clone https://github.com/AUTOMATIC1111/stable-diffusion-webui.git`



```
Command Prompt
Microsoft Windows [Version 10.0.19045.2965]
(c) Microsoft Corporation. All rights reserved.

C:\Users\Nabil Khaled>cd %userprofile%

C:\Users\Nabil Khaled>git clone https://github.com/AUTOMATIC1111/stable-diffusion-webui.git
```

After pressing enter, you should see an output indicating that a new folder, named 'stable-diffusion-webui', has been created in your 'home' directory. Keep in mind that you can save this to a different folder if you prefer, but if you do, make sure to use that new location in the subsequent steps.

Installation

Installing Git

```
Cloning into 'stable-diffusion-webui'...
remote: Enumerating objects: 12200, done.
remote: Counting objects: 100% (3/3), done.
remote: Compressing objects: 100% (3/3), done.
remote: Total 12200 (delta 0), reused 1 (delta 0), pack-reused 12197
Receiving objects: 100% (12200/12200), 24.20 MiB | 7.15 MiB/s, done.
Resolving deltas: 100% (8527/8527), done.
```

To move forward, you need to navigate to the new folder. You can type in either:

C:\Users\YOUR USER NAME\stable-diffusion-webui

or

%userprofile%\stable-diffusion-webui\models

and hit enter, the latter being simpler.

Within that folder, you'll find a .txt file named "Put Stable Diffusion checkpoints here.txt". Stay in this folder as you'll be placing some more folders here shortly.

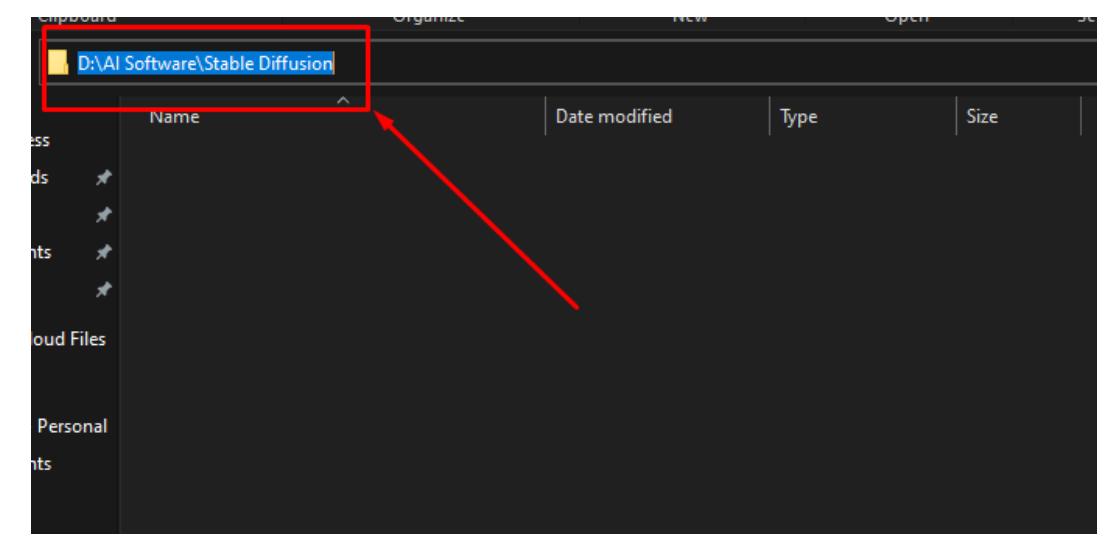
Note: To install Stable Diffusion in a different folder, press the Windows key + E to open your File Explorer. Navigate to the location where you'd like to install Stable Diffusion.

It's recommended to create a new folder, perhaps named Stable Diffusion or any name you prefer. Then, go to the address bar, copy, and paste the address, and use it as an alternative to the address we've been using in the previous and next steps (Refer to the following image).

Remember, Kevin Stratvert explains this process very well in his video for V1.5

<https://www.youtube.com/watch?v=onmqbl5XPH8>,

and Sebastian Kamph elucidates it for V2.1 at <https://www.youtube.com/watch?v=TZqo8kLmadE>.



Proceed to download Stable Diffusion v2.1, the latest version as of now, at this link: https://huggingface.co/stabilityai/stable-diffusion1-2/blob/main/v-768_1-2nonema-pruned.ckpt (or use the direct link: https://huggingface.co/stabilityai/stable-diffusion1-2/resolve/main/v-768_1-2nonema-pruned.ckpt).

If you wish to download an older version like V1.5, for educational reasons or otherwise, you can do so directly here: <https://huggingface.co/runwayml/stable-diffusion-v5-1/resolve/main/v-5-1pruned-emaonly.ckpt>. V1.5, which is my starting point with Stable Diffusion, was released on October 22. However, I suggest you go for V2.1, released on December 2022, 7 (direct link for V2.1: https://huggingface.co/stabilityai/stable-diffusion1-2/resolve/main/v-768_1-2nonema-pruned.ckpt).

After downloading your chosen Stable Diffusion version, move the downloaded file to your Stable Diffusion directory at (%userprofile%\stable-diffusion-webui), and place it in the "Models" folder.

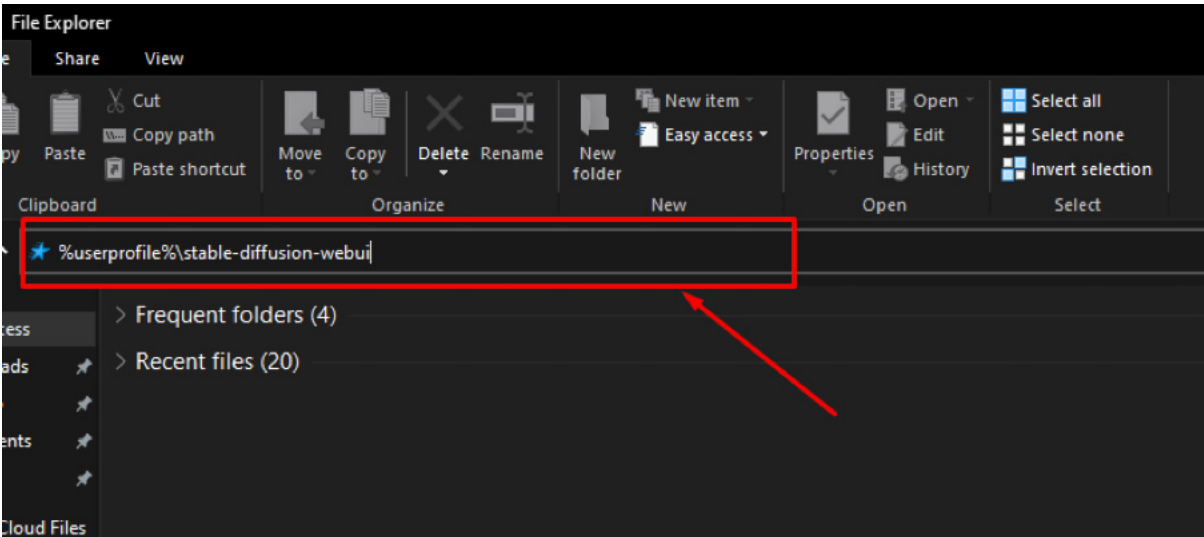
Installation

Installing Git

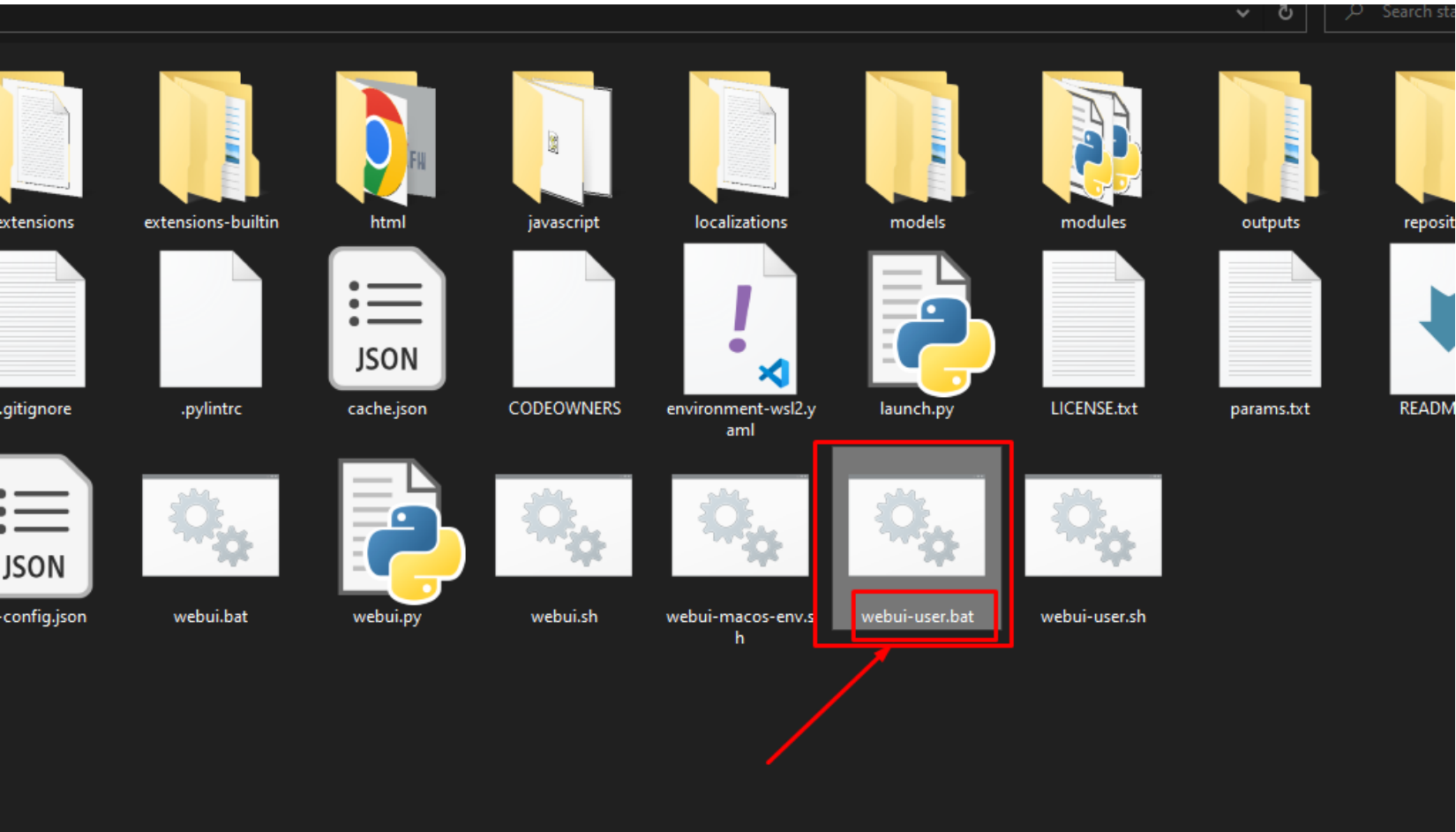
CRITICAL NOTE: If you're opting for V2.0 or V2.1, you'll need to do the following:

Go to this link: <https://raw.githubusercontent.com/Stability-AI/stablediffusion/main/configs/stable-diffusion/v-2inference-v.yaml>. Copy the text into a text file and save it as "v-768_1-2ema-pruned.yaml", or right-click on the page, choose "Save As", and use the same file name.

Now, you're ready to run the webui. You can find it in the same directory we visited earlier: (%userprofile%\stable-diffusion-webui). Just copy this link, paste it in your address bar, and hit enter (Refer to the following image).



Find a file called webui-user.bat. Double-click to run and complete the installation.



Installation

Installing Git

After running this batch file, let it work until it finishes, which might take a few minutes. Once it's done, you will see "Running on local URL: <http://127.0.0.1:7860>" on your screen.

Note: If you encounter any problems during this step, you can refer to discussions and solutions on Github, such as in this thread: <https://github.com/oobabooga/text-generation-webui/issues/985>.

Once it's completed, you can open your web browser and enter this URL: <http://127.0.0.1:7860/>. You will then be in the Stable Diffusion Environment.

In general, when you finish using Stable Diffusion, close the browser tab and the Command Prompt window (CMD). Be aware that if you close the CMD while using Stable Diffusion, it will cease to function since you've effectively shut it down.

If this wasn't convincing enough to understand why some people prefer to directly use websites like Hugging Face, the remaining details in this book will reveal more. It just takes a little effort to fully enjoy the immense power of this amazing image-generation AI tech.

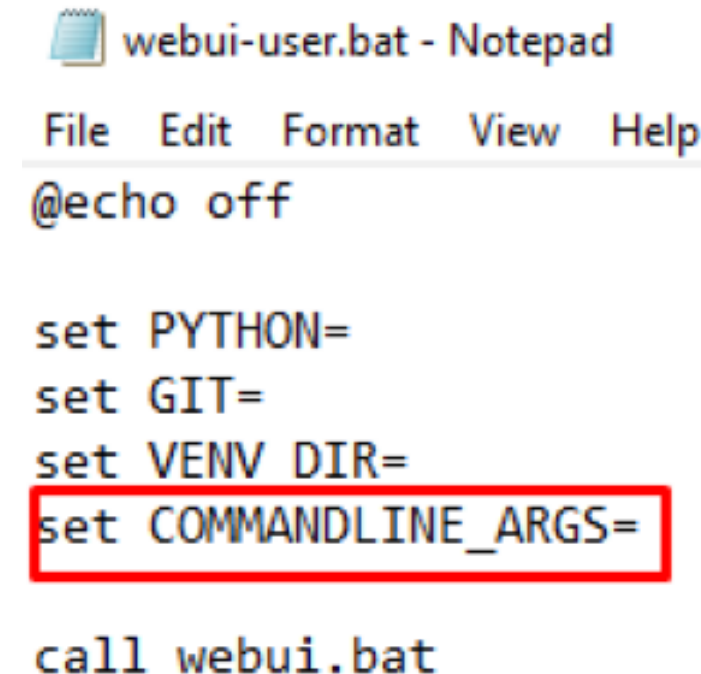
Remember, you can speed up Stable Diffusion with the "--xformers" option. If your GPU's VRAM is less than 8GB, it's recommended to enable the "--medvram" option. This will conserve memory, allowing you to generate more images simultaneously.

To enable these options, right-click on the "webui-user.bat" file and select "Edit". (You might need to select "Show More Options" first if you're using Windows 11.)

Replace the line
set COMMANDLINE_ARGS=

with
set COMMANDLINE_ARGS=--xformers --medvram

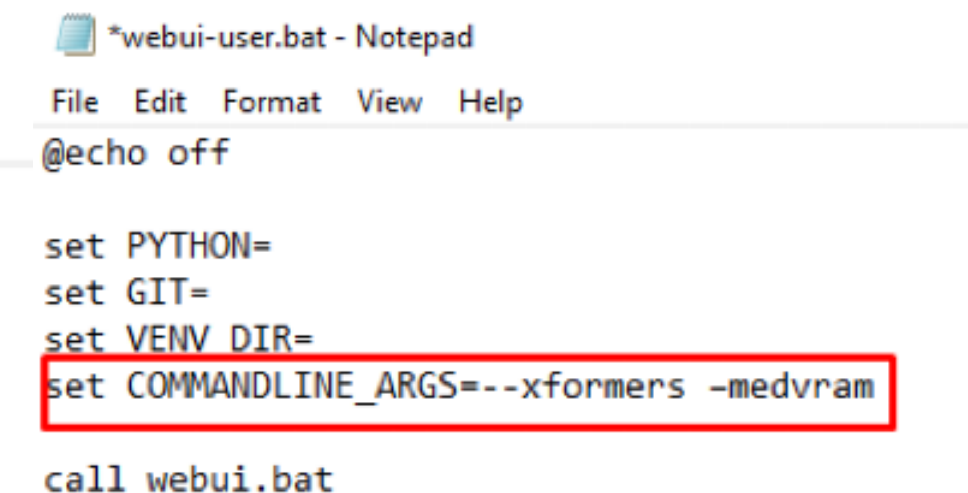
This modification will enable both the "--xformers" and "--medvram" options, thus optimizing the usage of Stable Diffusion on your system.



```
webui-user.bat - Notepad
File Edit Format View Help
@echo off

set PYTHON=
set GIT=
set VENV_DIR=
set COMMANDLINE_ARGS=

call webui.bat
```



```
*webui-user.bat - Notepad
File Edit Format View Help
@echo off

set PYTHON=
set GIT=
set VENV_DIR=
set COMMANDLINE_ARGS=--xformers --medvram

call webui.bat
```


Installation

Installing Git

When running after making this change you will see that it is installing xformers

```
git config --global --add safe.directory 'C:/Users/Nabil Khaled/stable-diffusion-webui'
venv "C:/Users/Nabil Khaled/stable-diffusion-webui/venv/Scripts/Python.exe"
Python 3.10.6 (tags/v3.10.6:9c7b4bd, Aug 1 2022, 21:53:49) [MSC v.1932 64 bit (AMD64)]
Commit hash: <none>
Installing xformers
```

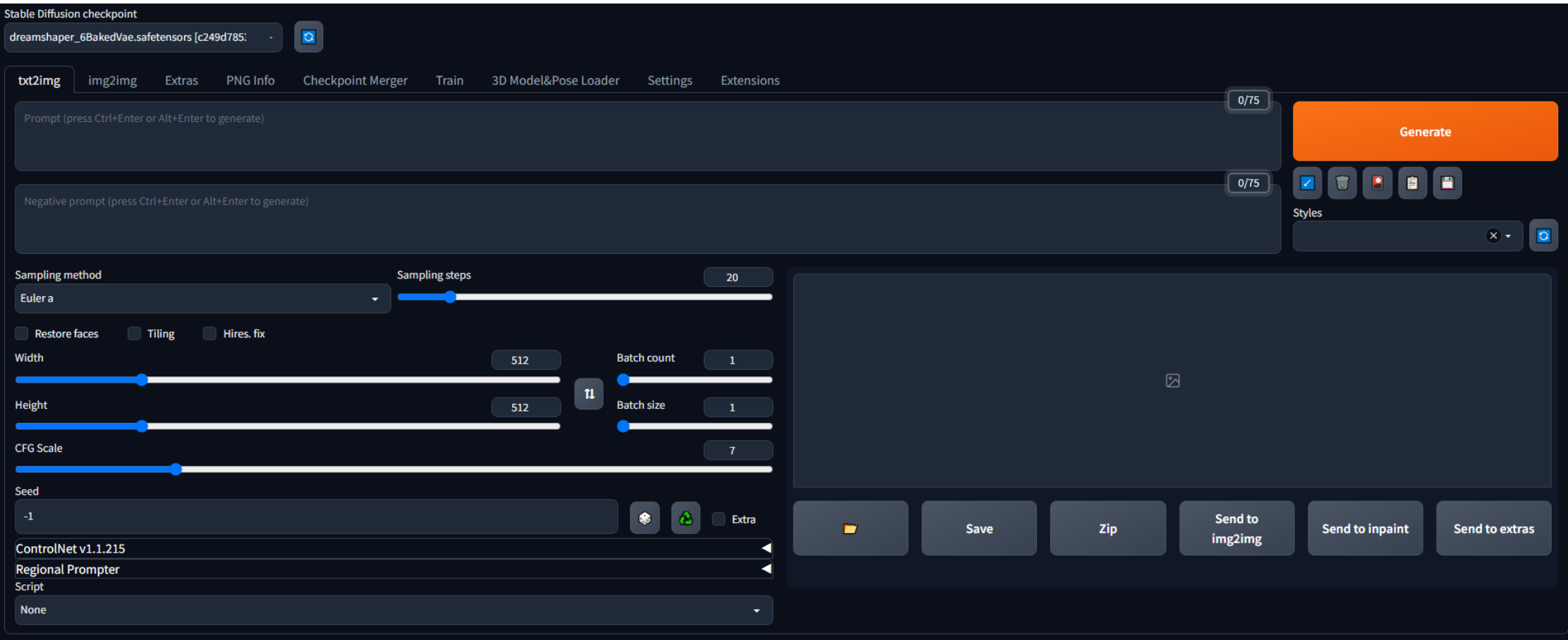
Important note:

If you're using a 4k series RTX graphics card, there are some additional steps that you need to follow to ensure proper installation and operation of Stable Diffusion. Here's the process as described on the Stable Diffusion Art website:

1. Navigate to the page <https://stable-diffusion-art.com/install-windows/>
 2. Follow the instructions specifically designed for 4k series RTX cards. They provide detailed steps and explanations to ensure the software functions correctly with your specific hardware.
- It's important to know that every GPU series might behave slightly differently due to architectural differences, driver support, and other factors. Thus, taking these additional steps ensures optimal compatibility and performance of Stable Diffusion with your 4k series RTX card.

Stable Diffusion Art is a highly respected and reliable source of information about Stable Diffusion, and they regularly update their content to reflect the latest changes, issues, and fixes related to the software. Their installation guide is very comprehensive and is regularly updated, making it an excellent resource for troubleshooting and learning more about Stable Diffusion.

Once you're done installing, as mentioned you can go to this URL in any browser and you will arrive inside the Stable Diffusion GUI. <http://127.0.0.1:7860/>



Installation

Installing Git

Updating AUTOMATIC1111 or any other software is crucial to keep it running smoothly, and to take advantage of new features and improvements. Here's how you can do it:

Updating AUTOMATIC1111 Every Time You Run It

If you want to ensure you always have the latest version of AUTOMATIC1111, you can modify the 'webui-user.bat' file to automatically update every time you run it. Here's how to do it:

Open 'webui-user.bat' file in a text editor.
Add the line 'git pull' just before the line 'call webui.bat'.
The file should now look like this:

```
@echo off

set PYTHON=
set GIT=
set VENV_DIR=
set COMMANDLINE_ARGS=--xformers --medvram
git pull
call webui.bat
```

This will ensure that AUTOMATIC1111 updates itself every time you start it.

Updating AUTOMATIC1111 As Needed

However, always using the latest version might not always be the best choice, as new versions might introduce unexpected issues or bugs. If you prefer to update AUTOMATIC1111 only when necessary, you can do so manually. Here's how:

Open Command Prompt (CMD).

Navigate to the directory where AUTOMATIC1111 is installed. If you installed it in the default location, you can use the command 'cd %userprofile%\stable-diffusion-webui'.

Once you're in the correct directory, run the command 'git pull'. This will update AUTOMATIC1111 to the latest version.

Remember to restart the software after updating it to make sure the new features and improvements are applied.

Below are potential challenges during Stable Diffusion installation and how to address them:

Windows Path Length Restrictions: Windows limits the length of file paths, which can cause issues when unzipping the downloaded file. To counter this, simply relocate the "stable-diffusion-ui" folder to the root level of your C: drive (or any other drive), such as C:\stable-diffusion-ui.

Absence of Miniconda3 Installer: Miniconda3, a lean Python distribution necessary for Stable Diffusion, may be missing. To install it, search "miniconda3" in your PC's Start Menu and follow the ensuing installation prompts.

Incorrect or Absent Python Version: The required Python version is crucial for Stable Diffusion to function. If you have an incompatible or missing Python version, download and install Python 3.10.6 from the official Python website. Make sure to add Python to the system PATH during installation and remove any previous versions if necessary.

Installation

Installing Git

Dependency Installation Errors: During installation, dependency-related issues like problems installing gfpGAN might arise. To resolve these, refer to the relevant documentation or community resources, and ensure all necessary dependencies are correctly installed.

Running Stable Diffusion on Different Systems: If you are trying to run Stable Diffusion on another machine after initial setup, you'll need to perform additional steps. Copy the entire Git clone, including dependencies for torchvision, to an external drive. Install Stable Diffusion on the new system, ensuring the same dependencies as the original system are set up. This promotes compatibility and smooth execution of Stable Diffusion.

If you would like alternative and easier ways to install Stable Diffusion, I would recommend trying these out:

Easy Diffusion V2.5, this is one of the easiest ways to install Stable Diffusion and it comes highly recommended, you can find the guide on GitHub here: <https://github.com/cmdr2/stable-diffusion-ui>

Installation

Resources

How to Install & Use Stable Diffusion on Windows
<https://www.youtube.com/watch?v=onmqbl5XPH8>

Sources of how to install on Windows:
<https://www.dexerto.com/tech/how-to-install-stable-diffusion2124809-/>
<https://stable-diffusion-art.com/install-windows/>

Kevin Stratvert who is a current popular tech youtuber an ex-Microsoft senior product manager with tons of insight explains how to install Stable Diffusion V1.5 here on his YouTube channel <https://www.youtube.com/watch?v=onmqbl5XPH8>

This tutorial by less popular but still intriguing Matt Wolfe is comprehensive and clear and we found it to be rather straightforward : <https://www.youtube.com/watch?v=Po-ykkCLE6M>

Although the videos out there on how to install V2.1 are less popular, you can find clear and concise explanations by YouTube channels such as Olivio Sarikas here

<https://www.youtube.com/watch?v=e3vcYVwEkW0>

Or an even more thorough explanation by Sebastian Kamph here: <https://www.youtube.com/watch?v=TZqo8kLmadE>

<https://www.youtube.com/watch?v=7AD6FFmrm04>

https://www.youtube.com/watch?v=ue-_BT0EtiY

Installation issue thread on GitHub: <https://github.com/CompVis/stable-diffusion/issues/484>

How to install Stable Diffusion V2.1:

<https://stable-diffusion-art.com/install-stable-diffusion1-2-/>

How to install Stable Diffusion V2.1 GUI on Automatic1111:
<https://aituts.com/install-stable-diffusion-v1-2/>

More about Easy Diffusion:
<https://stable-diffusion-ui.github.io/>

Stable Diffusion V2.0 No-code guide:
<https://medium.com/mlearning-ai/install-and-run-stable-diffusion-0-2-on-your-local-pc-no-code-guide-7e6270aaa9e>

07

REFERENCES

References

Intro & Comparison

Beginner's Guide to Stable Diffusion Art - Website: <https://stable-diffusion-art.com/beginners-guide/>

Stable Diffusion Models - Website: <https://stable-diffusion-art.com/models/>

Inpainting Basics in Stable Diffusion - Website: https://stable-diffusion-art.com/inpainting_basics/

Stable Diffusion - What, Why, How? - Video Link: <https://www.youtube.com/watch?v=ltLNyA3IWAQ&pp=ygUXZW1lcmdpbmcgdHJlbmRzIGFuZCB0ZWNoYm9sb2dpZXMgc3RhYmxlIGRpZmZ1c2lvbG3%D3%D>

Stable Diffusion vs Midjourney vs DALL-E2 - Link: <https://buildspace.so/notes/stable-diffusion-vs-midjourney-vs-dalle2->

Midjourney vs Stable Diffusion vs DALL-E: The Battle Royale of AI Image Models - Link: <https://medium.com/@neonforge/midjourney-vs-stable-diffusion-vs-dall-e-the-battle-royale-of-ai-image-models-comparison-test-ef50f6b4cbc8>

Digital Art Showdown: Stable Diffusion, DALL-E, and Midjourney - Link: <https://towardsdatascience.com/digital-art-showdown-stable-diffusion-dall-e-and-midjourney-db96d83d17cd>

Stable Diffusion vs Midjourney: AI Art Tools - Link: <https://www.digitaltrends.com/computing/stable-diffusion-vs-midjourney/>

DALL-E vs Midjourney AI vs Stable Diffusion: A Comparison of AI Models - Link: <https://blog.illacloud.com/dall-e-vs-midjourney-ai-vs-stable-diffusion-a-comparison-of-ai-models-that-can-generate-images-from-text/>

AI Image Generators Comparison - Part 1 - Video Link: <https://www.youtube.com/watch?v=dVI0rlqO41k&pp=ygUXZmZ1c2lvbiB2cyBtaWRqb3VybmV5IHZzIGRhGxl>

AI Image Generators Comparison - Part 2 - Video Link: <https://www.youtube.com/watch?v=KCj1HR7U9wA&pp=ygUXZmZ1c2lvbiB2cyBtaWRqb3VybmV5IHZzIGRhGxl>

Understanding Generation Models

Understanding Generation Models - Part 1 - Video Link: <https://www.youtube.com/watch?v=hflUstzHs9A&pp=ygUXZmZ1c2lvbiB2cyBtaWRqb3VybmV5IHZzIGRhGxl>

Understanding Generation Models - Part 2 - Video Link: <https://www.youtube.com/watch?v=yTAMrHVG1ew&pp=ygUXZmZ1c2lvbiB2cyBtaWRqb3VybmV5IHZzIGRhGxl>

Understanding Generation Models - Part 3 - Video Link: <https://www.youtube.com/watch?v=1ClpzeNxIhU>

LORA

LORA - Website: <https://stable-diffusion-art.com/LORA/>

LORA (Videos) - Video Link 1: <https://www.youtube.com/watch?v=ZHVdNeHZ>

LORA (Videos) - Video Link 1: <https://www.youtube.com/watch?v=ZHVdNeHZPdc&pp=ygUXZmZ1c2lvbiB2cyBtaWRqb3VybmV5IHZzIGRhGxl>

Olivio Sarikas on LORAs: <https://www.youtube.com/watch?v=9MT1n97ITaE>

Aitreprenuer on LORA training: <https://www.youtube.com/watch?v=70H03cv-57o>

References

Stable Diffusion Videos

Protogen Model (Video) - Video Link: <https://www.youtube.com/watch?v=HAR6LjzTg5k>

Install Stable Diffusion Locally (Quick Setup Guide) (Video) - Video Link: <https://www.youtube.com/watch?v=Po-ykkCLE6M&pp=ygUhC2V0dGluZyB1cCBzdGFibGUgZGlmZnVzaW9uIHdlYnVp>

Local Stable Diffusion Settings: (EXPLAINED!!) (Video) - Video Link: <https://www.youtube.com/watch?v=Z3IHmdqUar0>

How To Install Stable Diffusion With Prompting Cheat Sheets (Video) - Video Link: <https://www.youtube.com/watch?v=xMrilkJ21yo>

Stable Diffusion Got Supercharged - For Free! (Video) - Video Link: <https://www.youtube.com/watch?v=1RvZWHtFXuY>

The RIGHT Way to Create INFINITE AI Animations | Deform Tutorial (Video) - Video Link: <https://www.youtube.com/watch?v=bicPayZDI60>

Better Than Midjourney: Openjourney Stable Diffusion (Video) - Video Link: https://www.youtube.com/watch?v=gLk_DCmLFYM

Stable Diffusion Tutorial: ULTIMATE guide - everything you need to know! (Video) - Video Link: <https://www.youtube.com/watch?v=DHaL56P6f5M>

UNCENSORED GPT4 x Alpaca Beats GPT 4! Create ANY Character! - Video Link: <https://www.youtube.com/watch?v=nVC9D9fRyNU>

ControlNet (Video) - Video Link: <https://www.youtube.com/watch?v=rCygkyMuSQo>

CFG (Video) - Video Link: <https://www.youtube.com/watch?v=kuhO9zAzetk&pp=ygUeQ0ZHIHBhcmFtZXRIciBzdGFibGUgZGlmZnVzaW9u>

Local Stable Diffusion Settings: <https://www.youtube.com/watch?v=Z3IHmdqUar0>

CFG (Video) - Video Link: <https://www.youtube.com/watch?v=kuhO9zAzetk&pp=ygUeQ0ZHIHBhcmFtZXRIciBzdGFibGUgZGlmZnVzaW9u>

Video Link: <https://www.youtube.com/watch?v=5QTcuk2kcwA&pp=ygUeQ0ZHIHBhcmFtZXRIciBzdGFibGUgZGlmZnVzaW9u>

ControlNet

ControlNet - Website: <https://stable-diffusion-art.com/ControlNet/>

ControlNet (GitHub) - Link: <https://github.com/lllyasviel/ControlNet>

ControlNet Animation (Video) - Video Link: <https://www.youtube.com/watch?v=EAXUInT70TA>

QR Codes with ControlNet - Video Link: <https://www.youtube.com/watch?v=OJLD0SIB7nl>

QR Code with ControlNet 1.1: <https://www.youtube.com/watch?v=IntRn96C4l4>

Sebastian Kamph - ControlNet Workflow: <https://www.youtube.com/watch?v=4u-Ytioi3DM>

ControlNet 1.1 by Olivio Sarikas: <https://www.youtube.com/watch?v=zrGLEgGFJY4>

ControlNet by Kamph: <https://www.youtube.com/watch?v=tBwmbTwMxfQ>

References

Outpainting

Outpainting - Website: <https://stable-diffusion-art.com/outpainting/>

How to Use Outpainting (Blog) - Link: <https://blog.opendream.ai/how-to-use-outpainting>

Negative Prompts

Writing Prompts - Negative Prompts (Wiki) - Link: <https://github.com/cmdr2/stable-diffusion-ui/wiki/Writing-prompts#negative-prompts>

Cheat Sheets

Stable Diffusion Cheat Sheet - Link: <https://supagruen.github.io/StableDiffusion-CheatSheet/>

Stable Diffusion Cheat Sheet - Link: https://docs.google.com/spreadsheets/d/1SRqJ7F_6yHVS0eCi3U82aA448TqEGrUIRrLLZ51abLg/htmlview#

Stable Diffusion Cheat Sheet - Link: https://reentry.org/artists_sd-v1-4

Stable Diffusion Cheat Sheet - Link: <https://www.urania.ai/top-sd-artists>

Stable Diffusion Cheat Sheet - Link: <https://stablediffusion.fr/artists>

Stable Diffusion Cheat Sheet - Link: <https://sdartists.app/#/list>

Stable Diffusion Cheat Sheet - Link: <https://sgreens.notion.site/sgreens/4ca6f4e229e24da6845b6d49e6b08ae7?v=fd861d1c65d456e98904fe3f3670bd3>

Stable Diffusion Cheat Sheet - Link: <https://proximacentaurib.notion.site/e28a4f8d97724f14a784>

Stable Diffusion Cheat Sheet - Link: <https://proximacentaurib.notion.site/e28a4f8d97724f14a784a538b8589e7d?v=ab624266c6a44413b42a6c57a41d828c>

Stable Diffusion Cheat Sheet - Link: <https://dict.latentspace.observer/>

Stable Diffusion Cheat Sheet - Link: <http://xosh.org/clip-artists/>

Stable Diffusion Cheat Sheet - Link: <https://www.the-ai-art.com/modifiers>

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Artists and Models

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