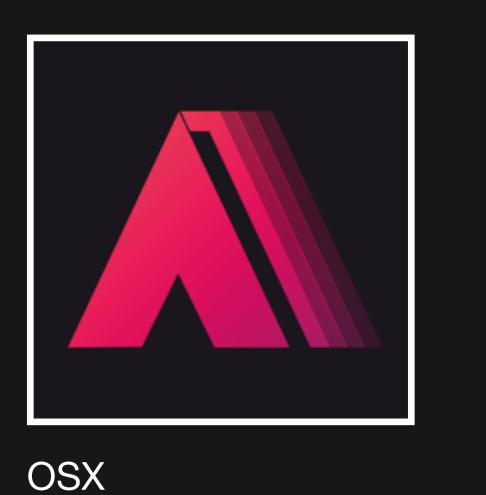


LET'S GO // WebUi

- > Stable Diffusion icon on the desktop
- Python (console)
- WebUi (browser)

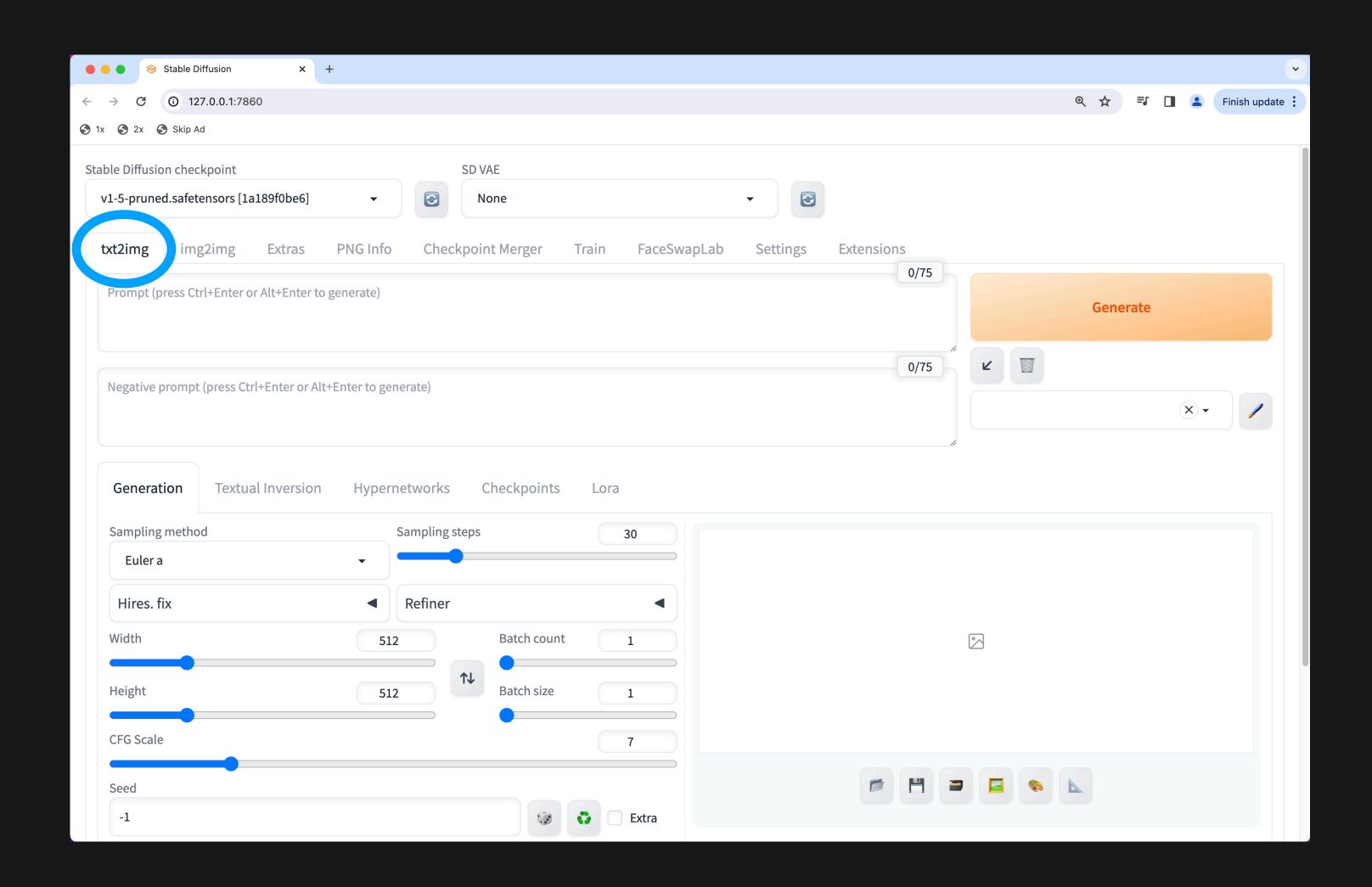




TEXT TO IMAGE // txt2img

Generate images from text prompts.

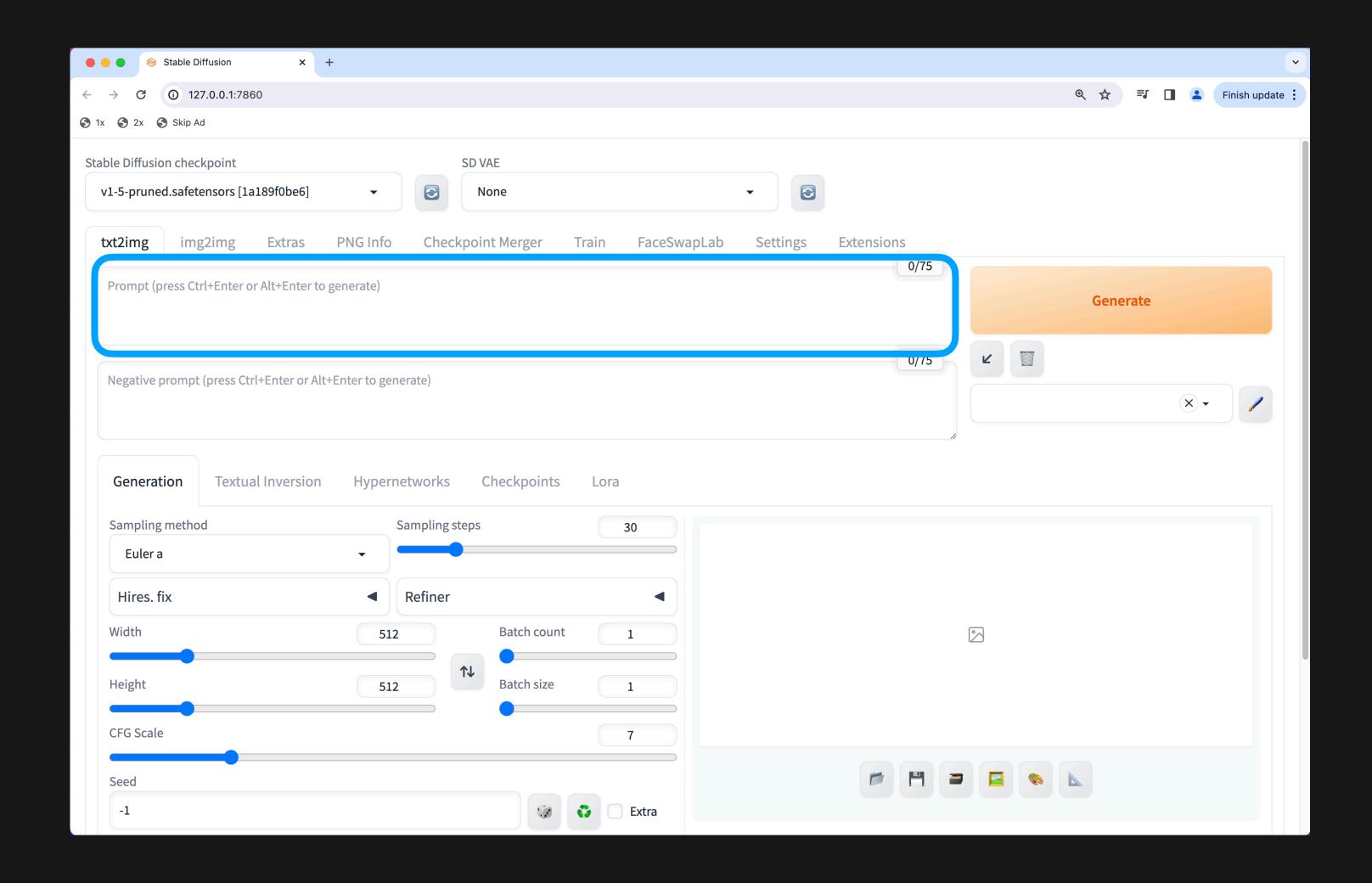
> select txt2img tab



PROMPT (simple)

Describe the image you want to generate.

- > Try a simple prompt first
- > "A dog playing poker"
- > Hit GENERATE

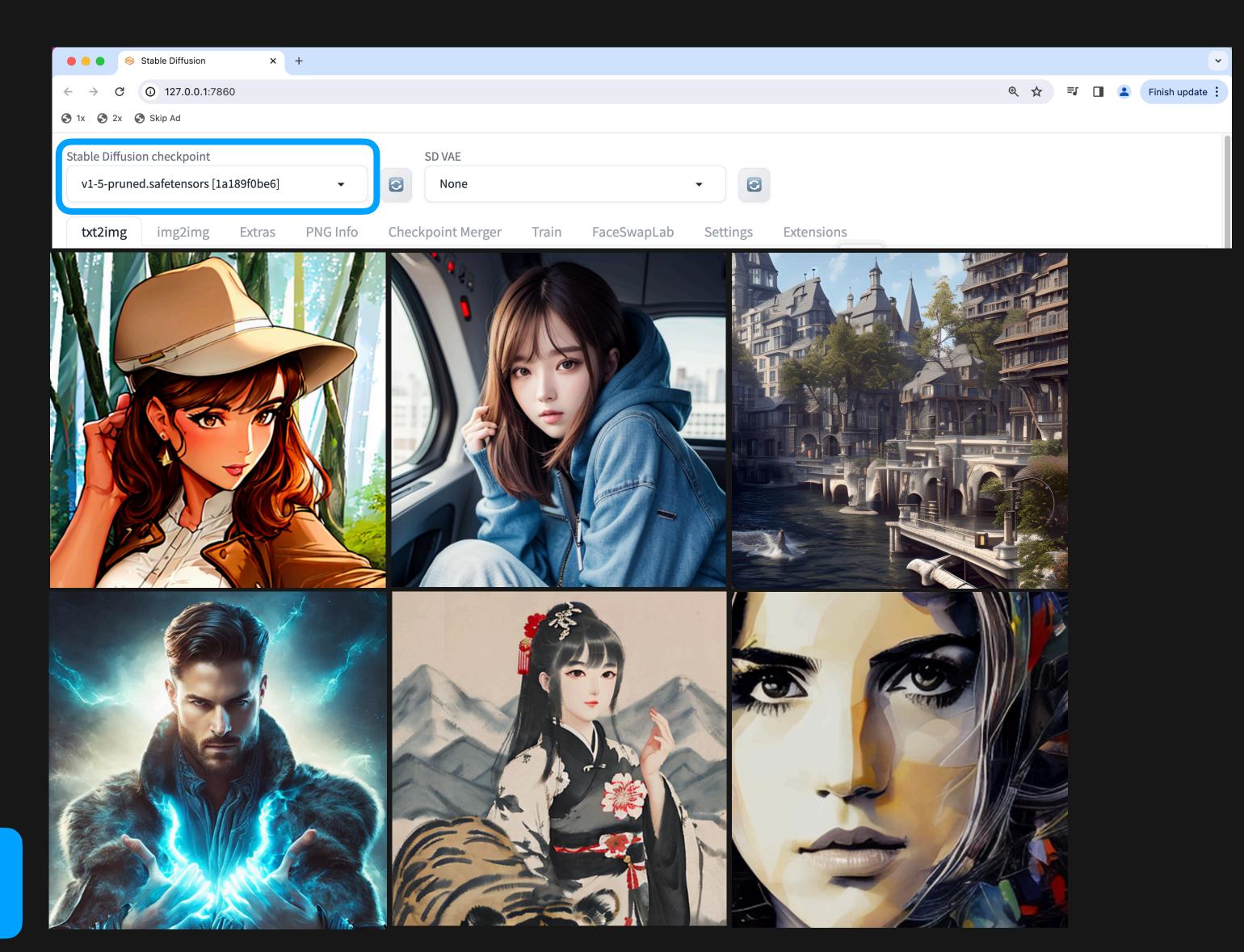


CHECKPOINT MODELS

Diffusion weights / Models trained on a specific style.

- > illustrations
- > paintings
- > anime
- > fantasy
- > landscapes
- > animals
- > fashion styles
- > cgi / realistic paintings
- > photorealism
- > (...)

Midjourney does NOT have models & you have to adjust your style/mediaum via prompt!



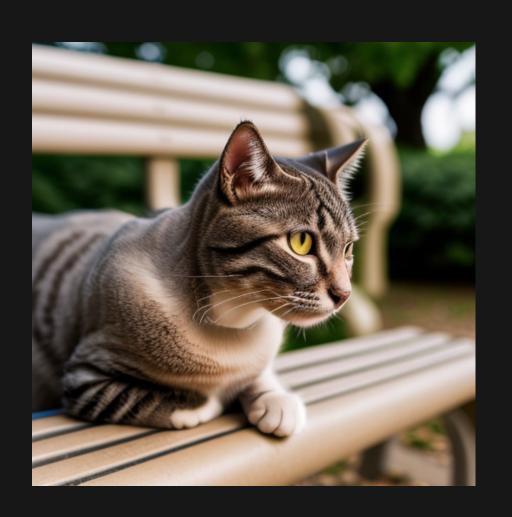
CHECKPOINT MODELS

"A cat on a bench"

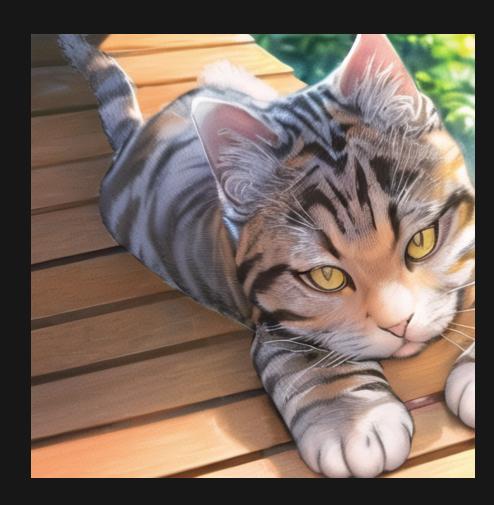
v1.5-pruned (Base Model)



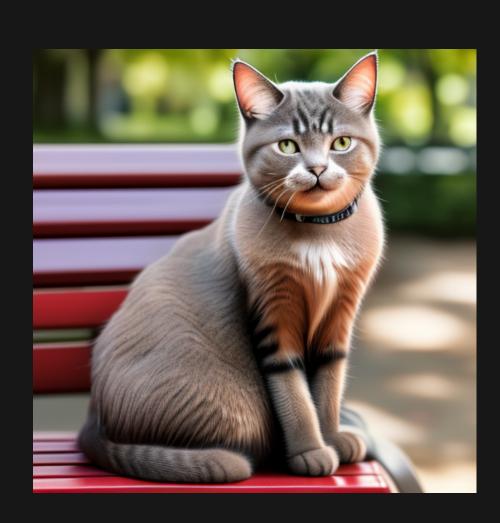
RealisticVision (photoreal)



Anything-V3 (Anime)



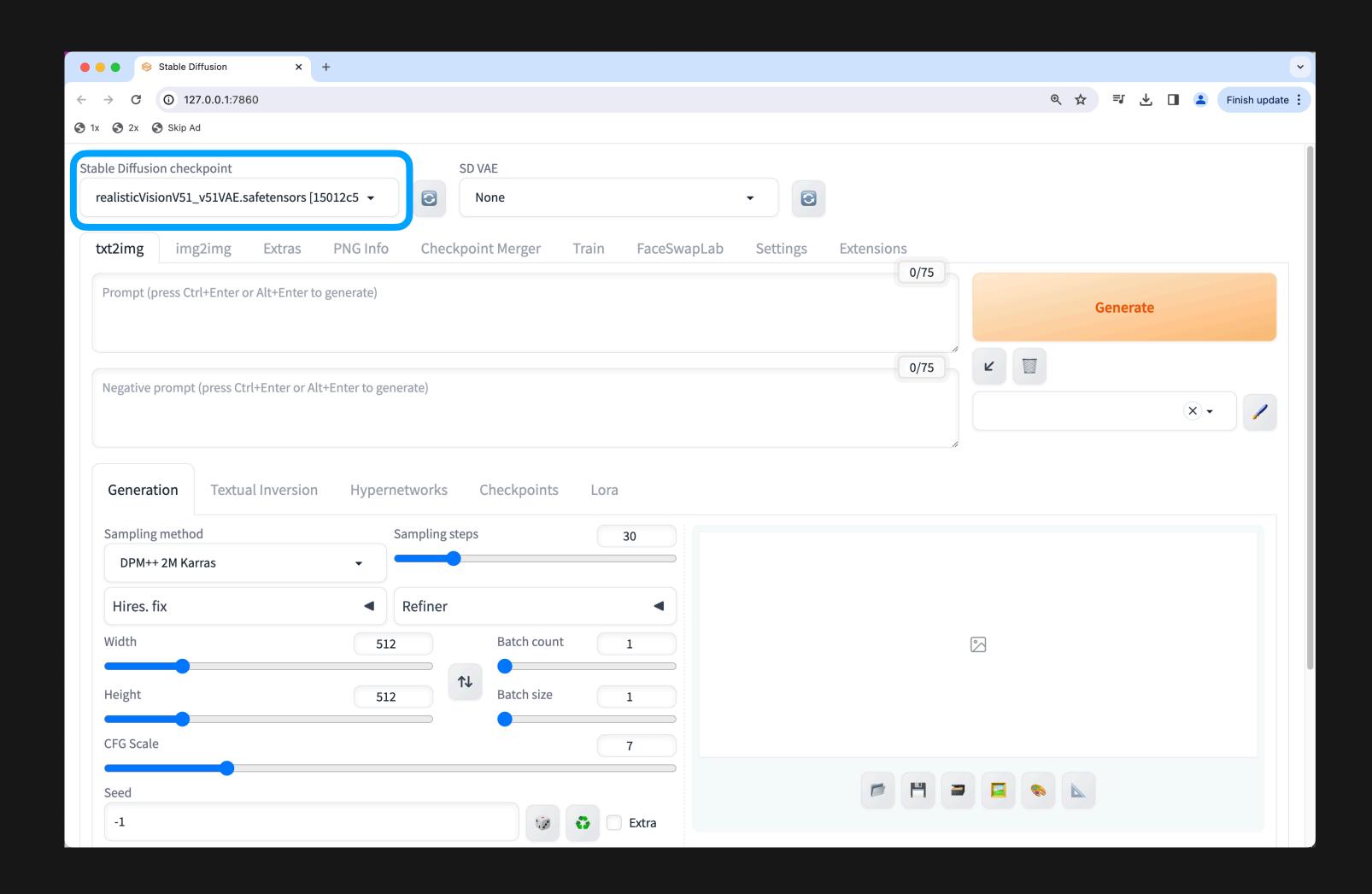
Deliberate-V3 (CGI)



CHECKPOINT MODELS

Same prompt with different models ... Let's try it out!

- Anything-V3 (Anime)
- RealisticVision (photo-real)



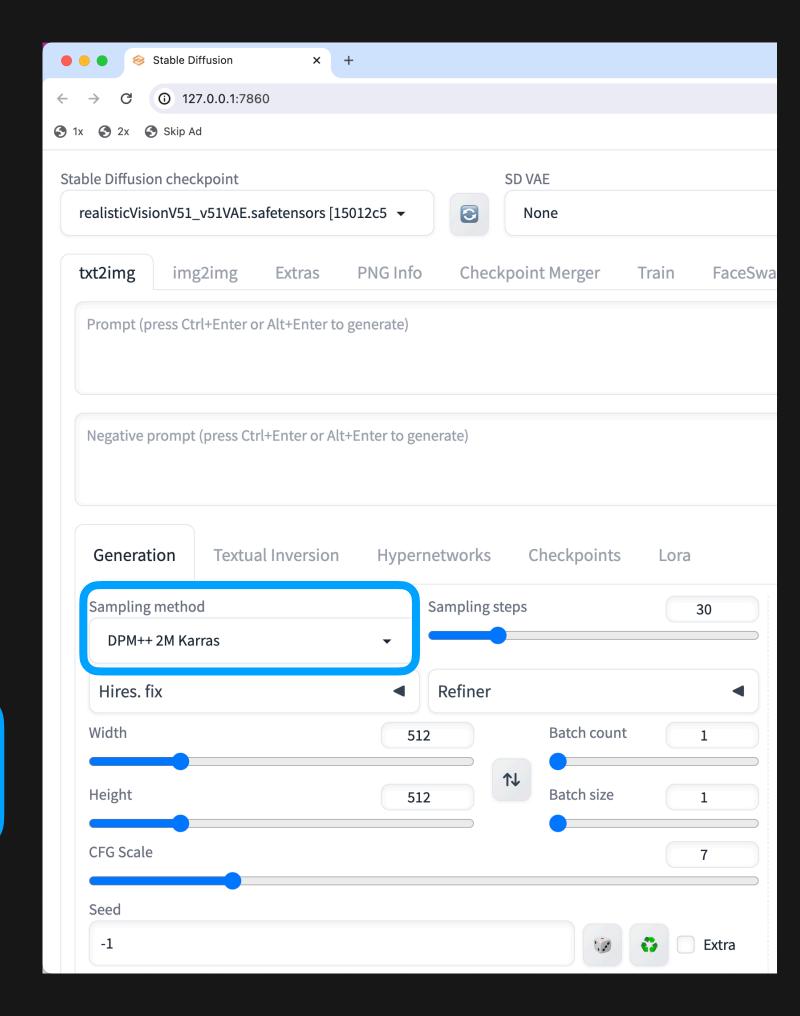
SETTINGS: Sampling Methods

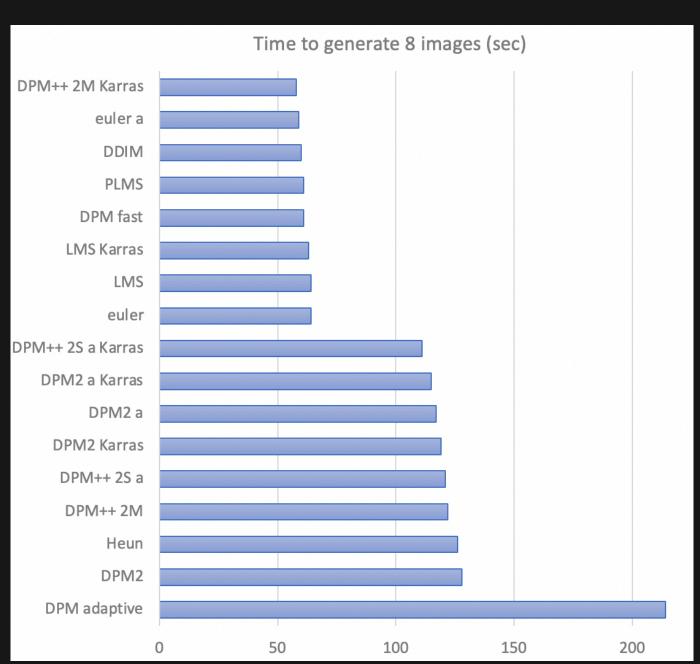
Sampling Methods (model dependent)

Euler a: Fast

Heun: Detailed for big GPUs

DPM++ 2M Karras: Detailed



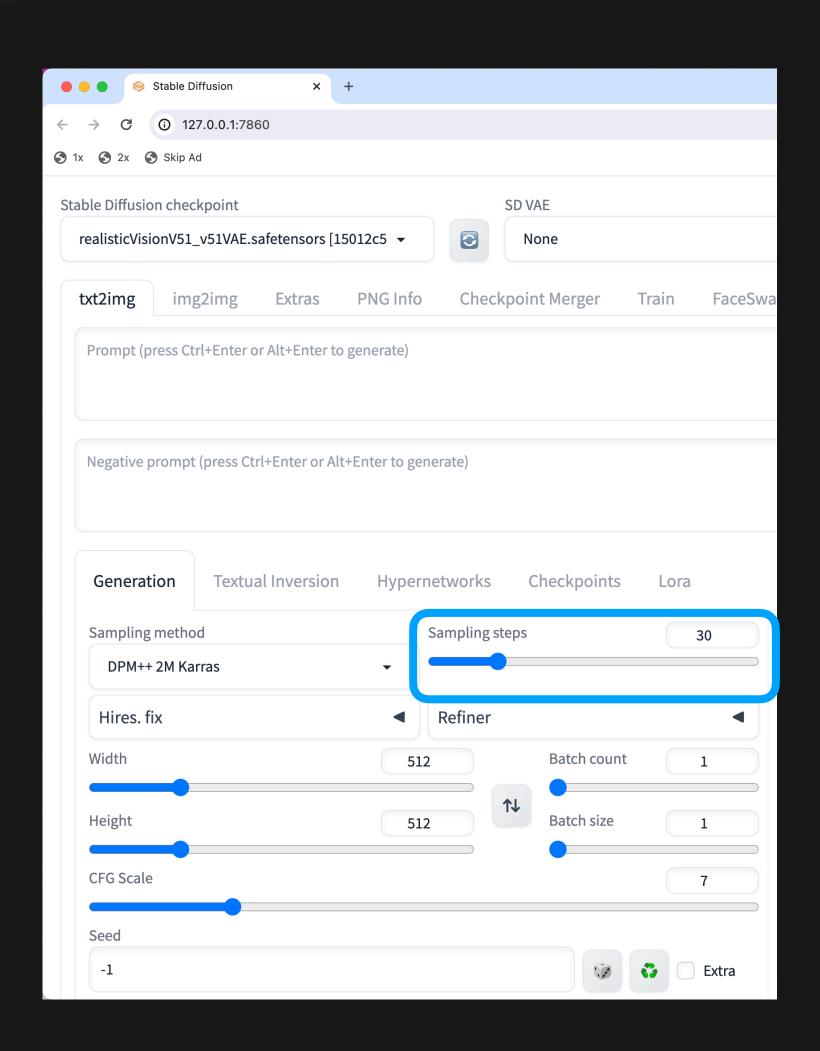


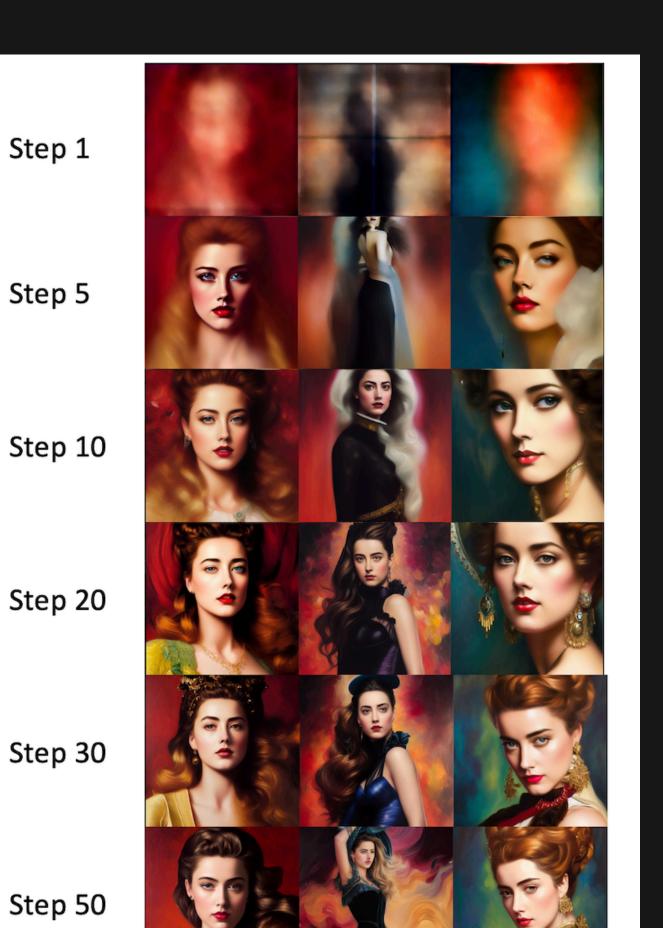
SETTINGS: Sampling Steps

Sampling Steps (Level of detail / Iterations)

Low number: Fast & rough High number: Slow & detailed

Recommended: 20 - 35 Steps





SETTINGS: Image Size

Size of output image.

 $512 \times 512 = gold standard$

512 x 768 (portrait) = OK

768 x 512 (landscape) = OK

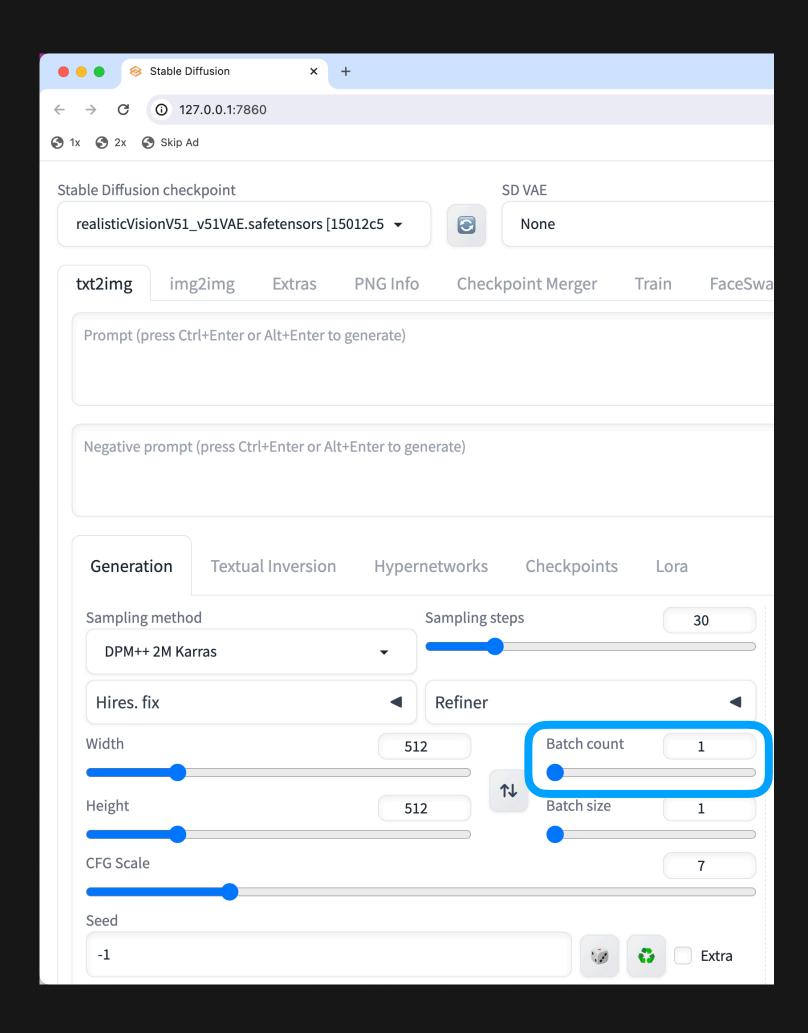
Stable D	iffusion x -	+					
→ C ① 12	7.0.0.1:7860						
1x 🔇 2x 🔇 Skip A	d						
Stable Diffusion checkpoint				SD VAE			
realisticVisionV51_	v51VAE.safetensors [15	012c5 ▼	N	lone			
txt2img img	g2img Extras	PNG Info	Checkpoi	nt Merger	Train	FaceSw	
Prompt (press Ct	rl+Enter or Alt+Enter to	generate)					
Negative prompt	(press Ctrl+Enter or Alt-	+Enter to genera	ate)				
	To 1 1 1 1 2 2 2 2 2 2				1		
Generation	Textual Inversion	Hypernet	Works	Checkpoints	Lora		
Sampling method		Sampling steps				30	
DPM++ 2M Ka	rras	•					
Hires. fix		■ R	tefiner			•	
Width		512		Batch count		1	
Height		512	↑	Batch size		1	
		312					
CFG Scale						7	
Cood							
Seed -1						Extra	
_						LALIA	

SETTINGS: Batch Size

Batch count = Number of images generated each time (in series = takes more time)

Batch size = Number of images generated in parallel (To save time = don't touch it)

Recommendation:
Set batch count to
3 or 4 or more...

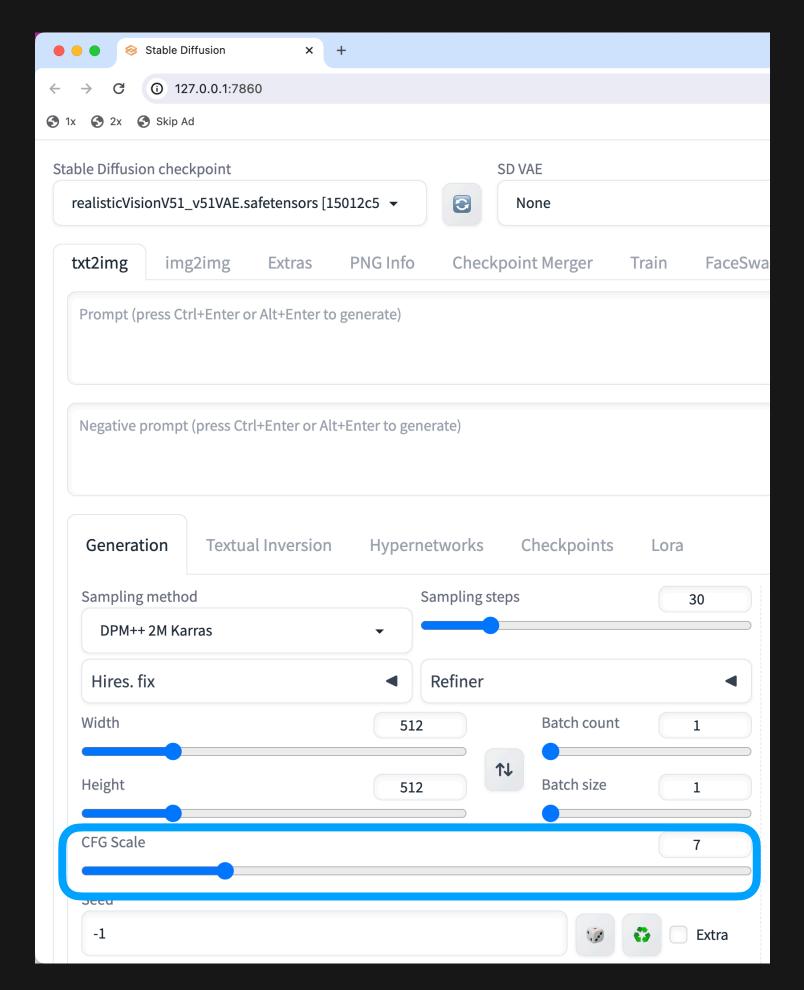


SETTINGS: CFG Scale

CFG = parameter to control How much the model should respect your prompt.

- 1 mostly ignore prompt
- 3 be creative with prompt
- 7 Good balance
- 15 Adhere more to prompt
- 30 strictly follow prompt

Recommendation: 7 Increase to 10 for more control Stay away from extremes.



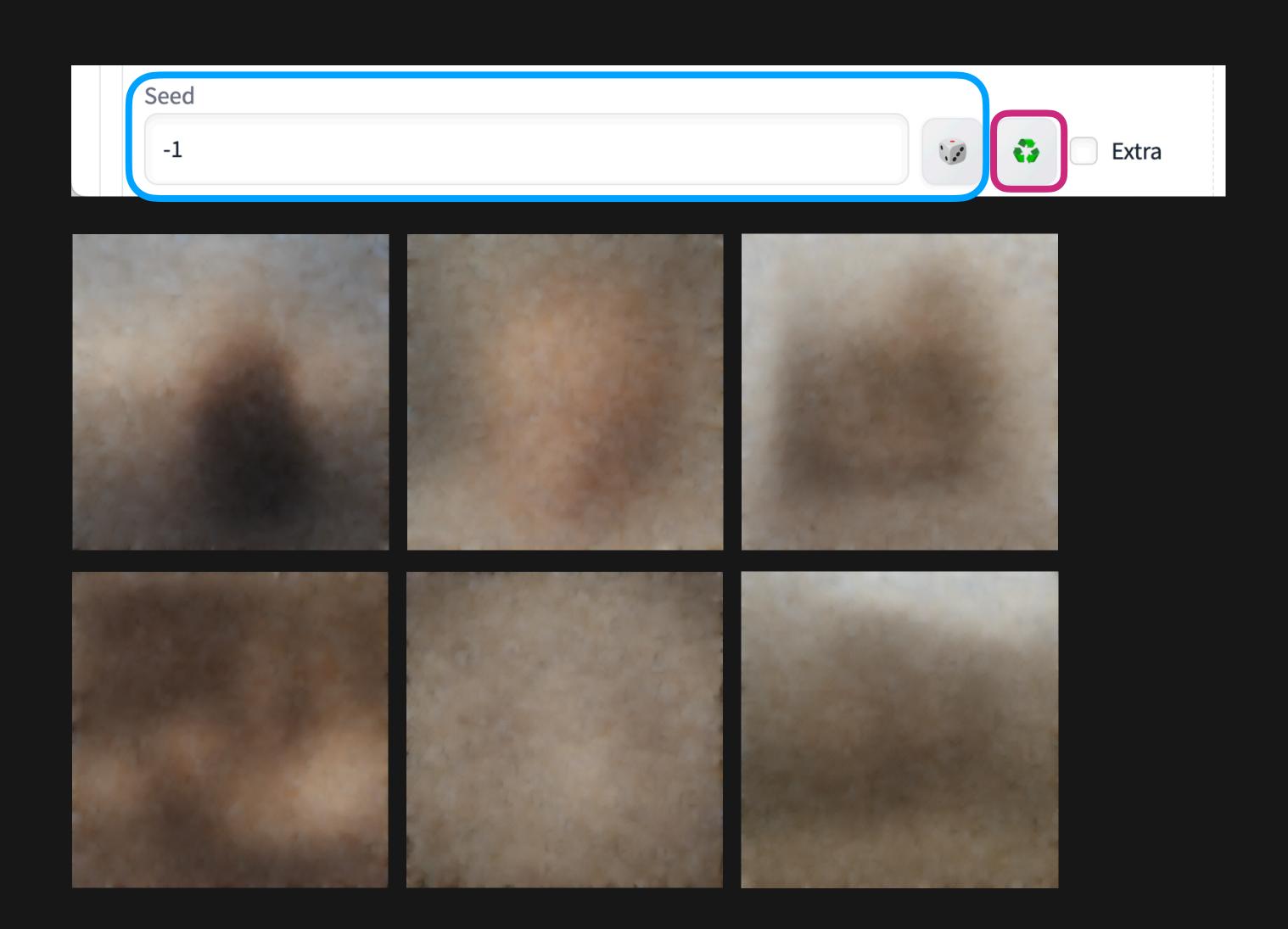


SETTINGS: SEED

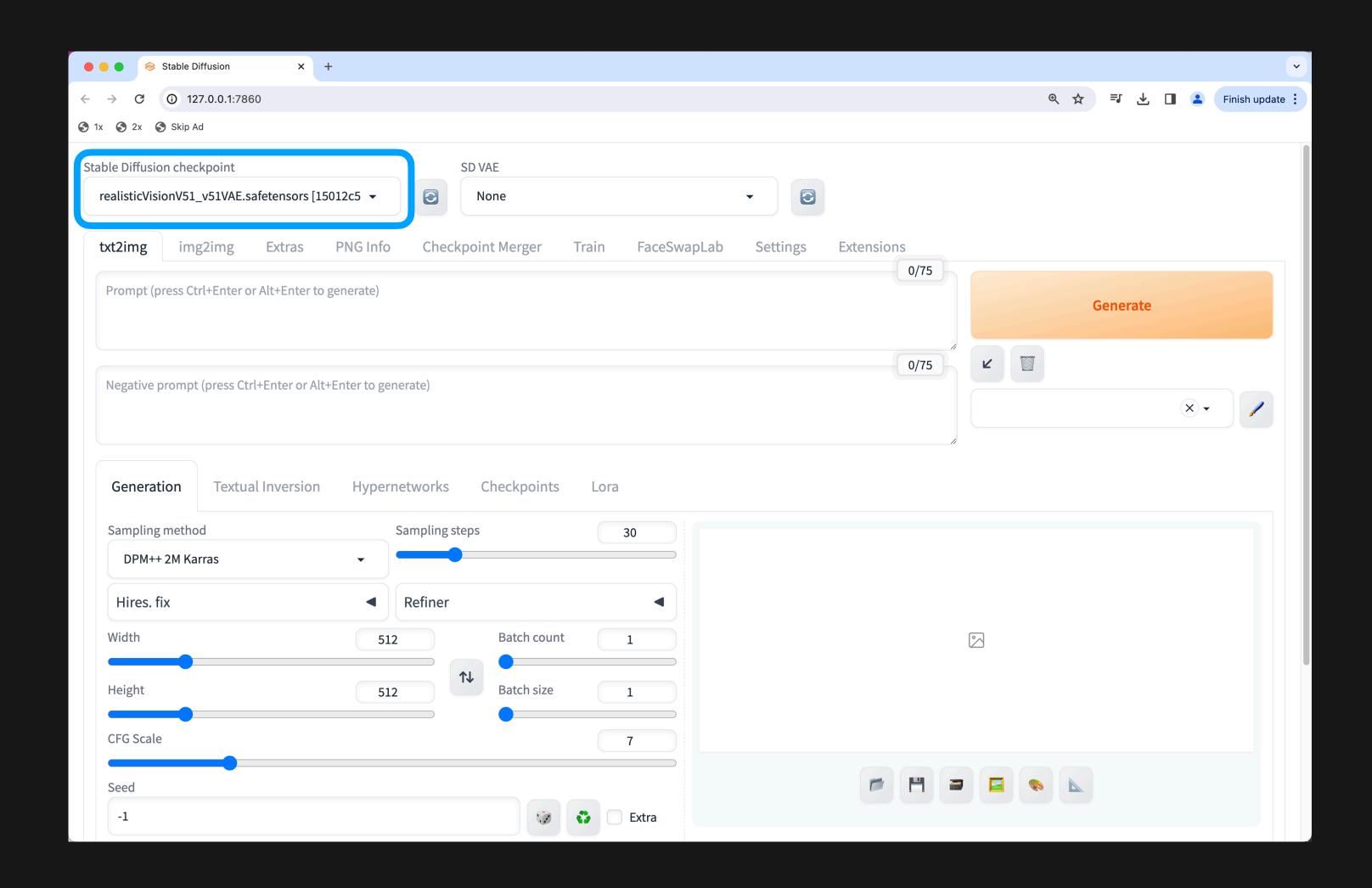
Seed "-1" = random noise used to initialise the generation process. Useful for generating new pictures with the same prompt.

Seeds can also be re-used.

Useful for making changes to the prompt.



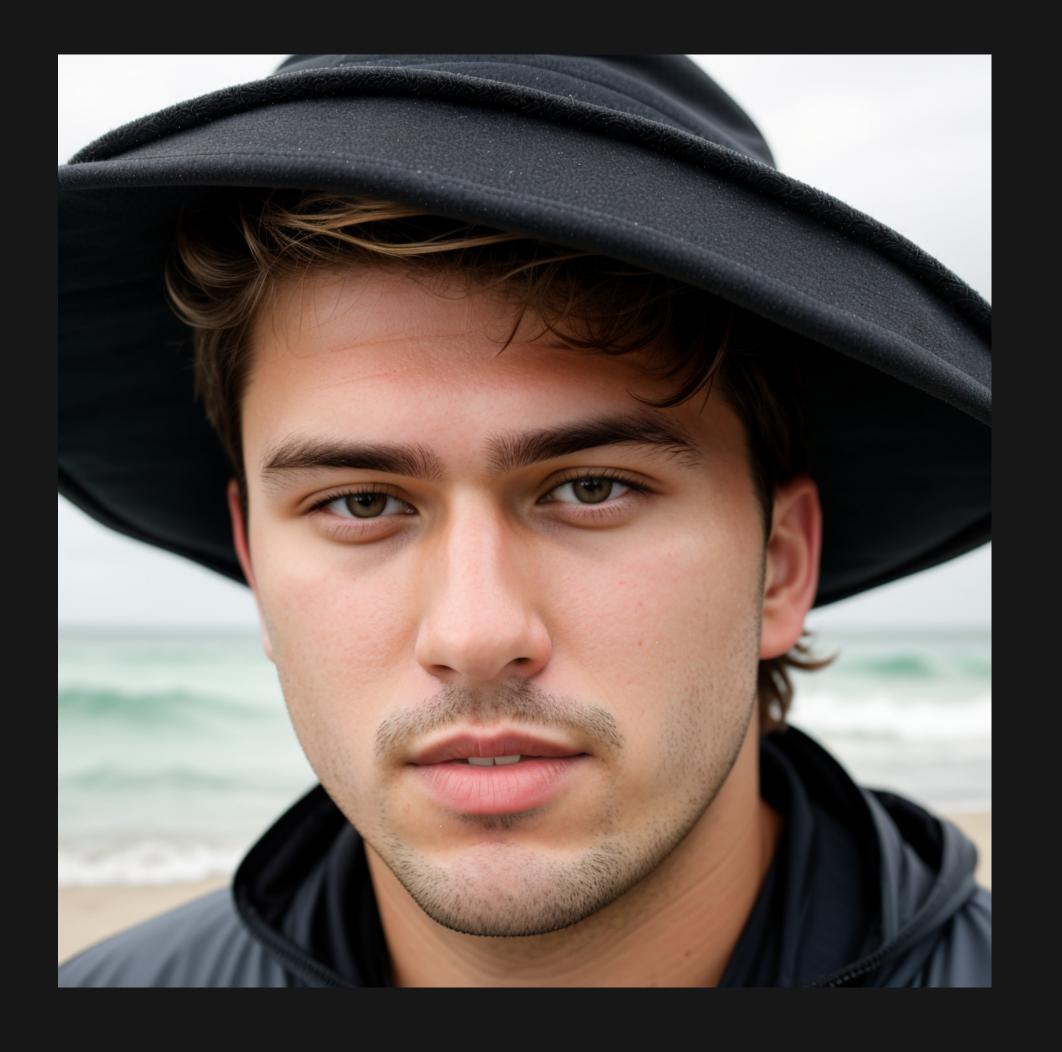
Let's all use these settings and go for a photorealistic image ...



Goal:

A man at the sea.

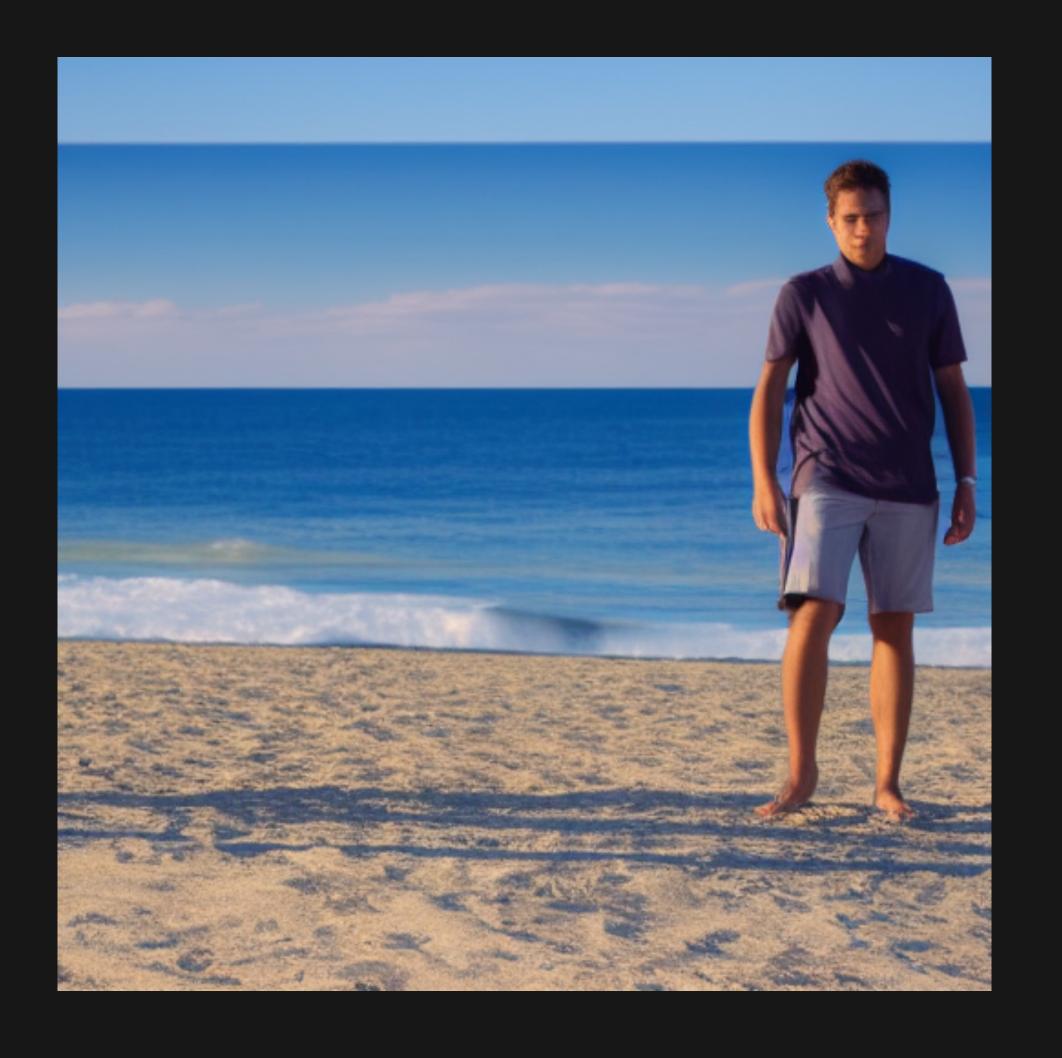
(...)



Outcome:

A man at the sea.

> What went wrong ???



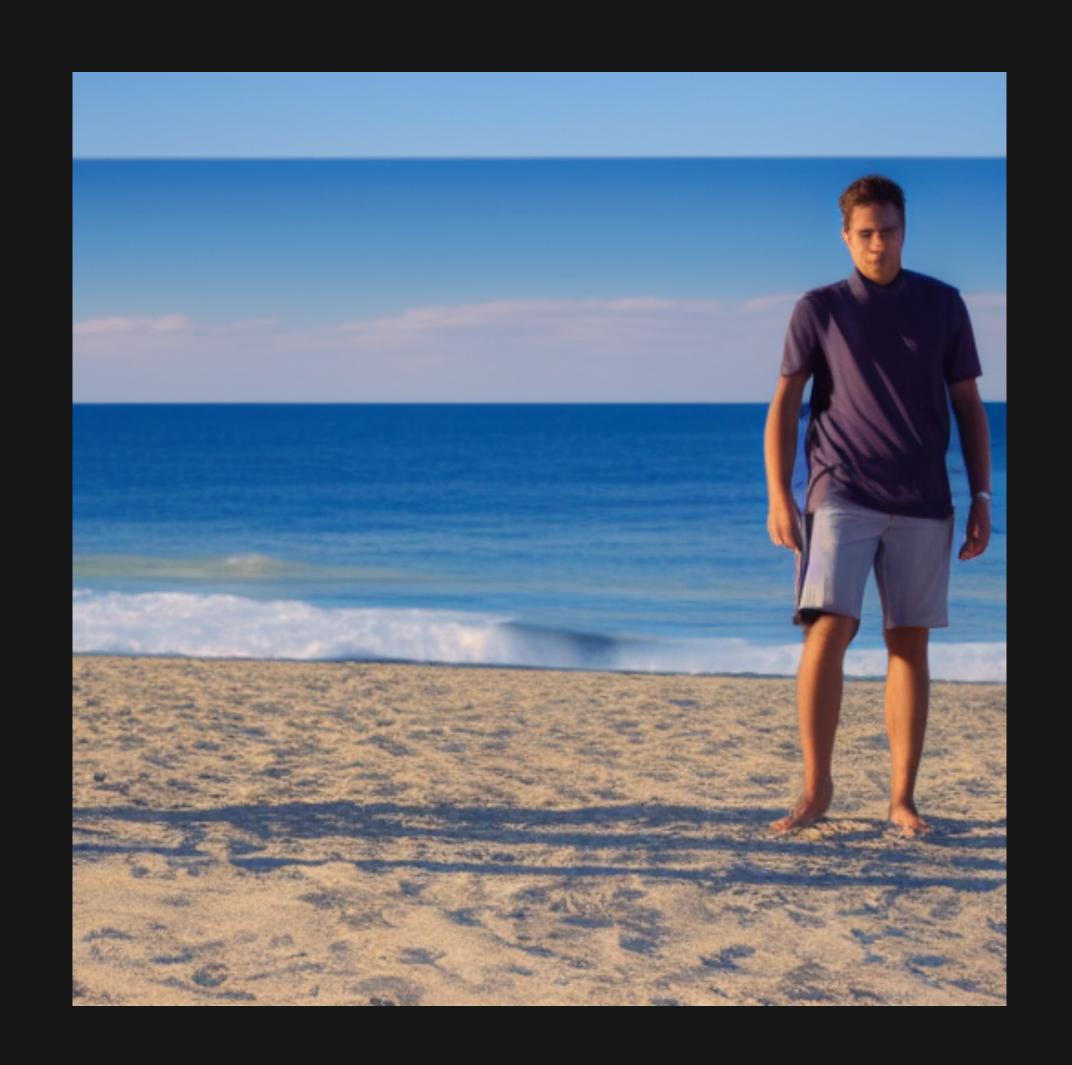
Outcome:

A man at the sea.

> What went wrong ???

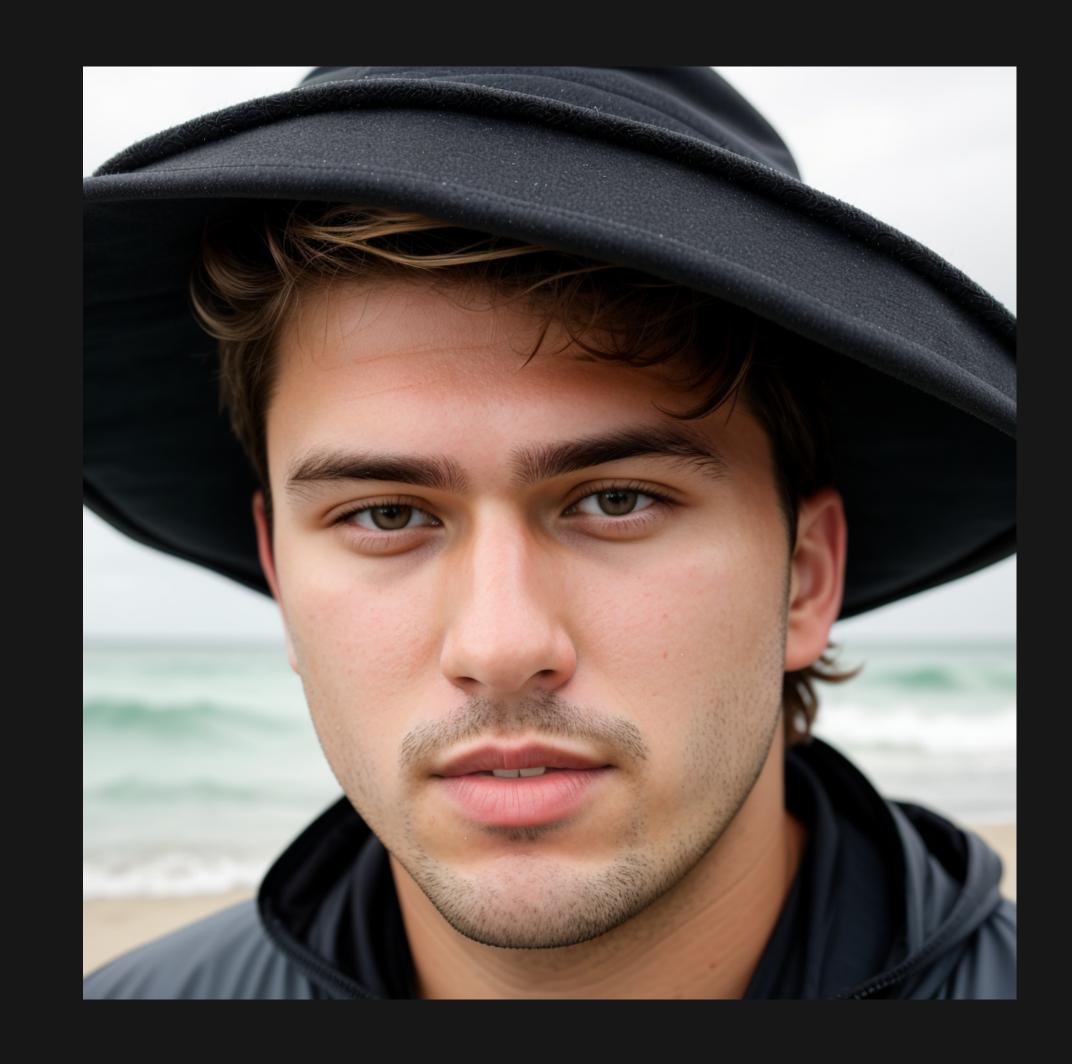
Prompts need to be clear and concise. Describe the image in enough detail.

Start with a specific subject in mind and add keywords to steer towards a particular effect.



Anatomy of a good prompt:

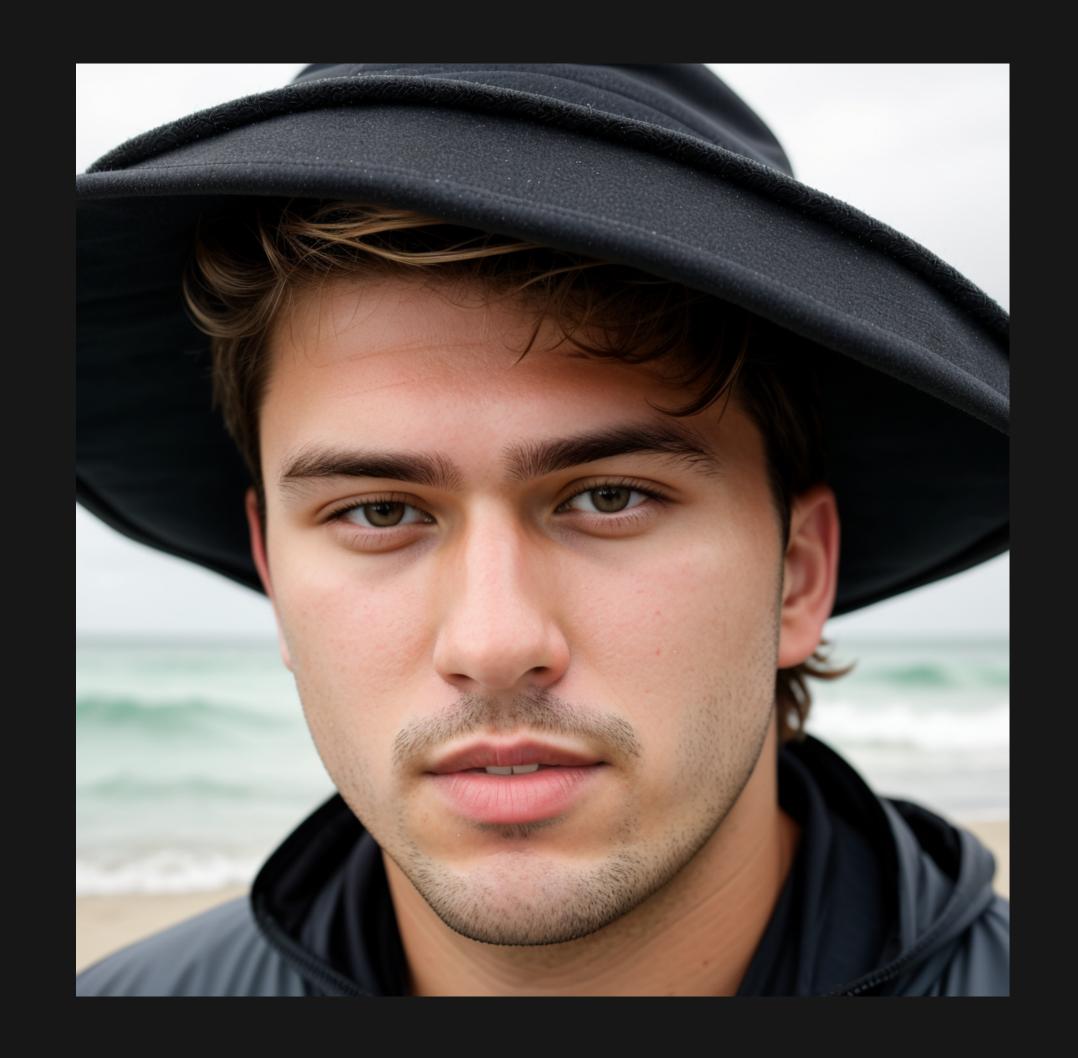
- 1. Subject
- 2. Medium
- 3. Style
- 4. Artist / Website
- 5. Resolution
- 6. Additional Details
- 7. Color
- 8. Lighting



1. Subject

A young caucasian man wearing black clothes.

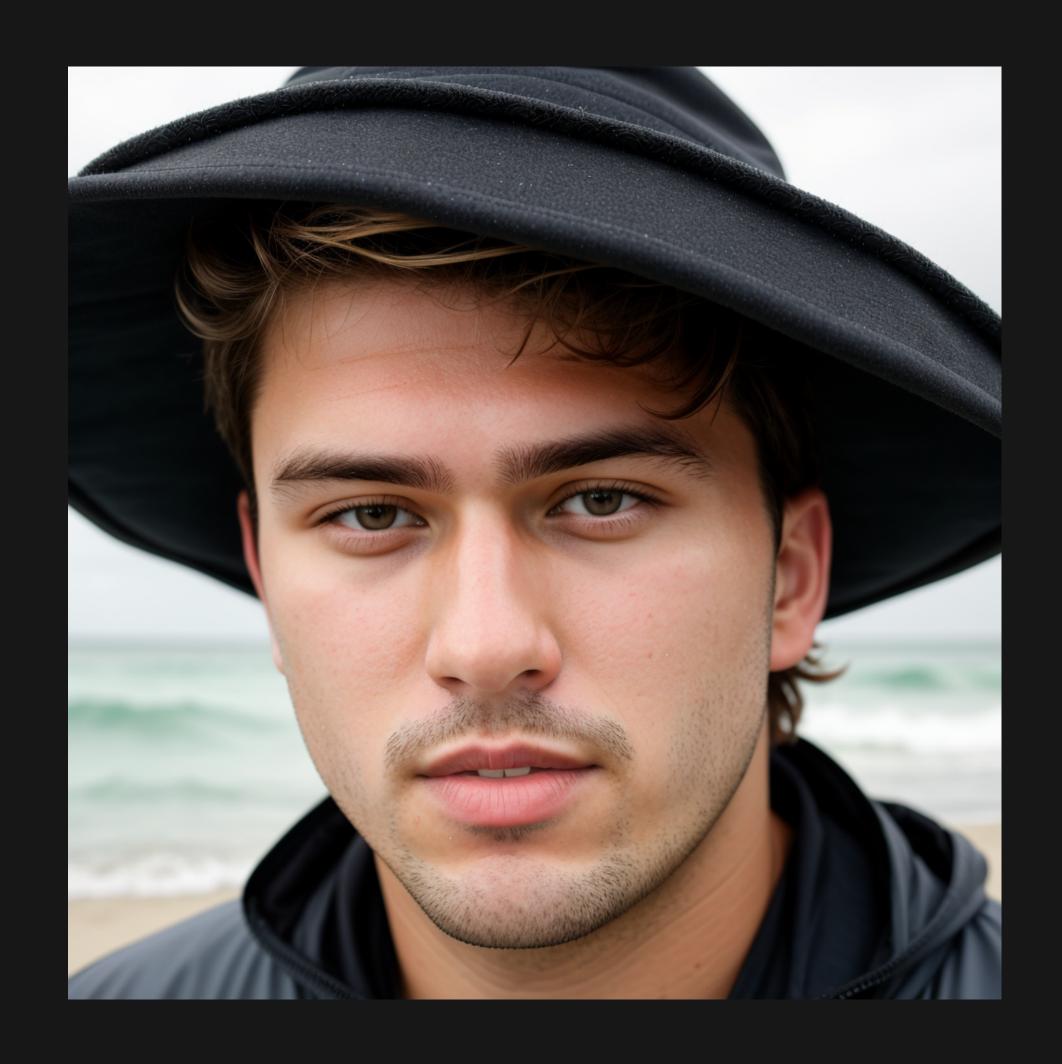
What is young? > 28 y.o. caucasian man wearing black clothes



2. Medium

Digital painting, illustration, watercolour painting, ink drawing, etc.

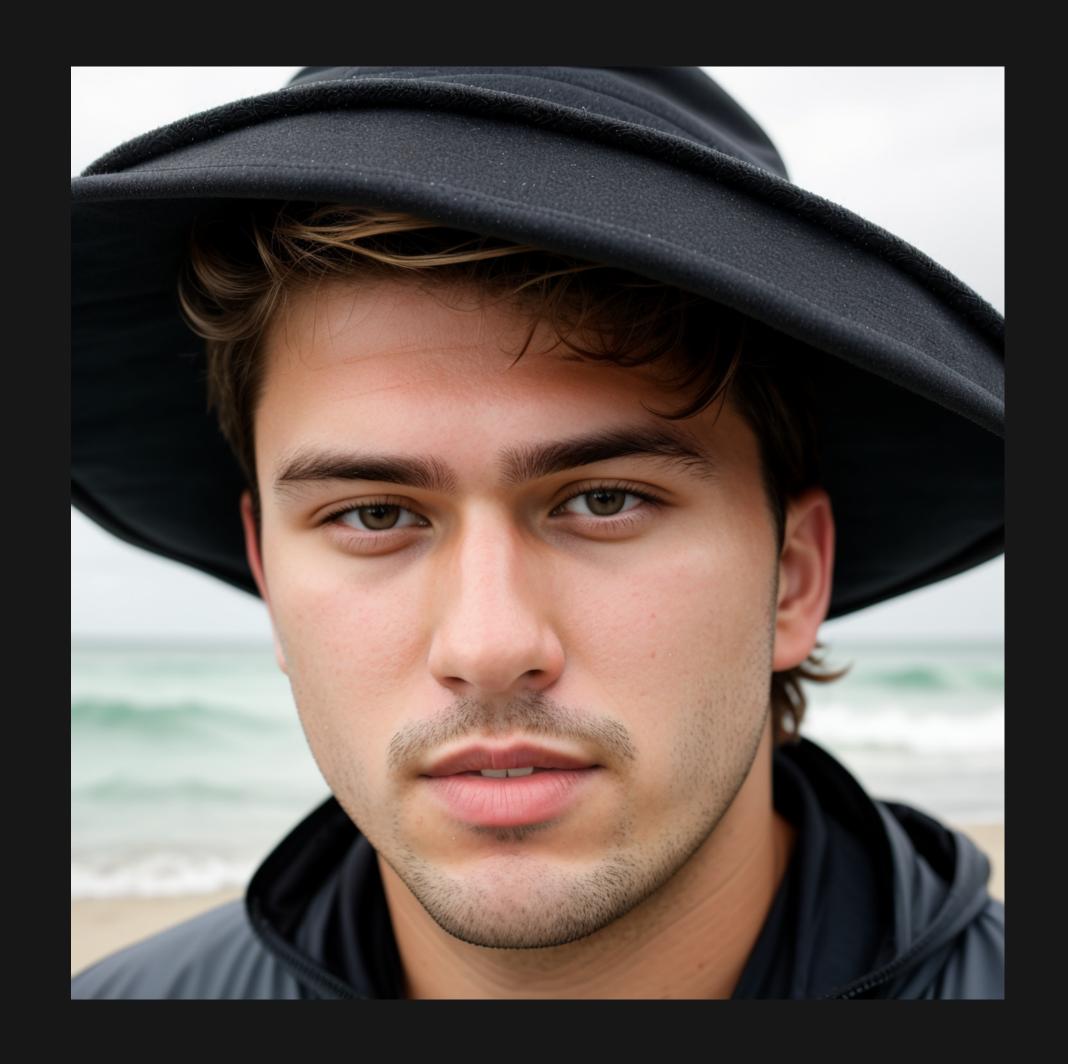
> RAW photo



3. Style

Hyperrealistic, fantasy, impressionist, surrealist, pop art, etc.

> close-up, dslr, film grain, Fujifilm XT3

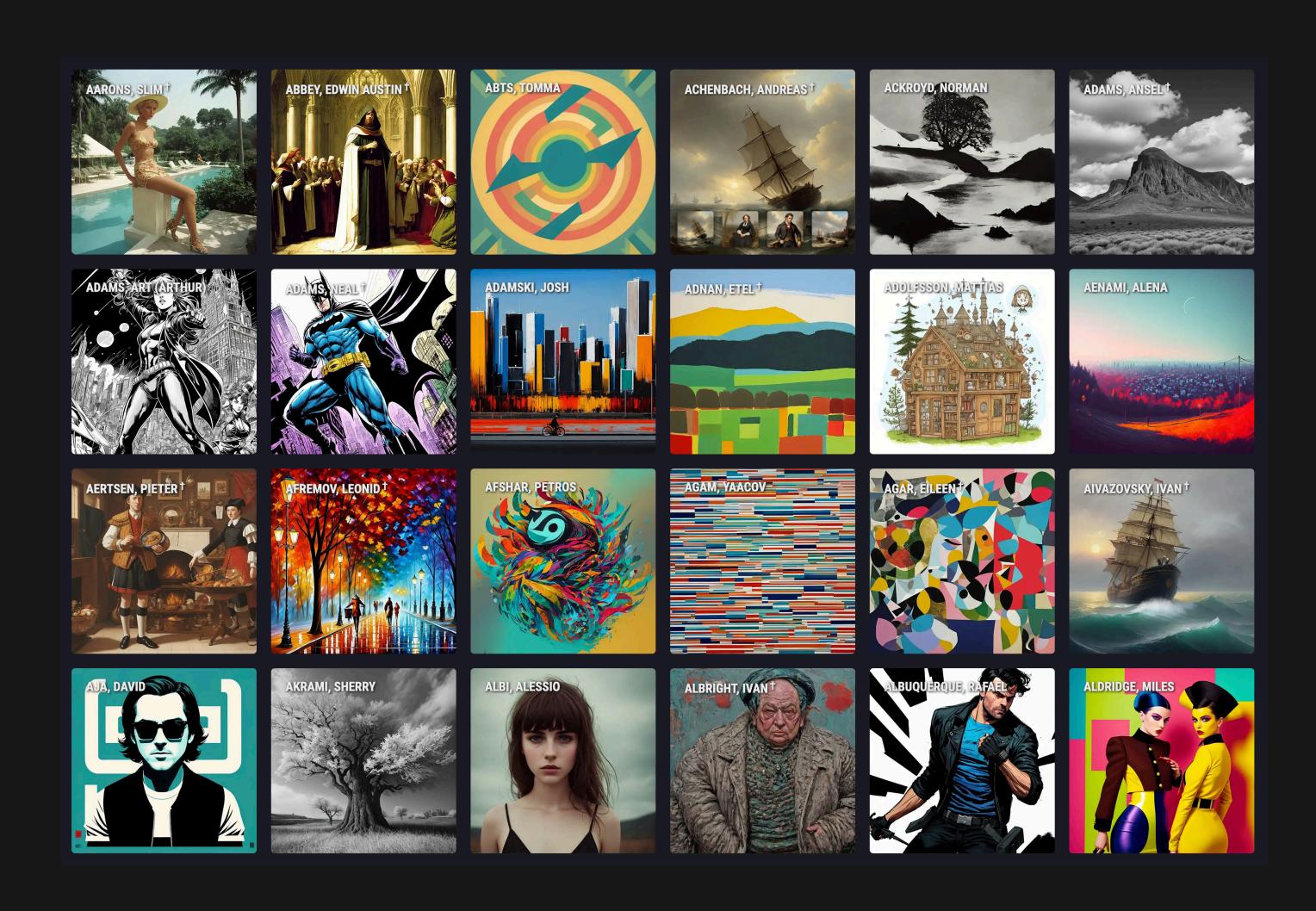


4. Artist / Website

https://supagruen.github.io/StableDiffusion-CheatSheet/

Roy Lichtenstein, Paul Zezanne, Claude Monet, Artstation, Deviant Art, etc.

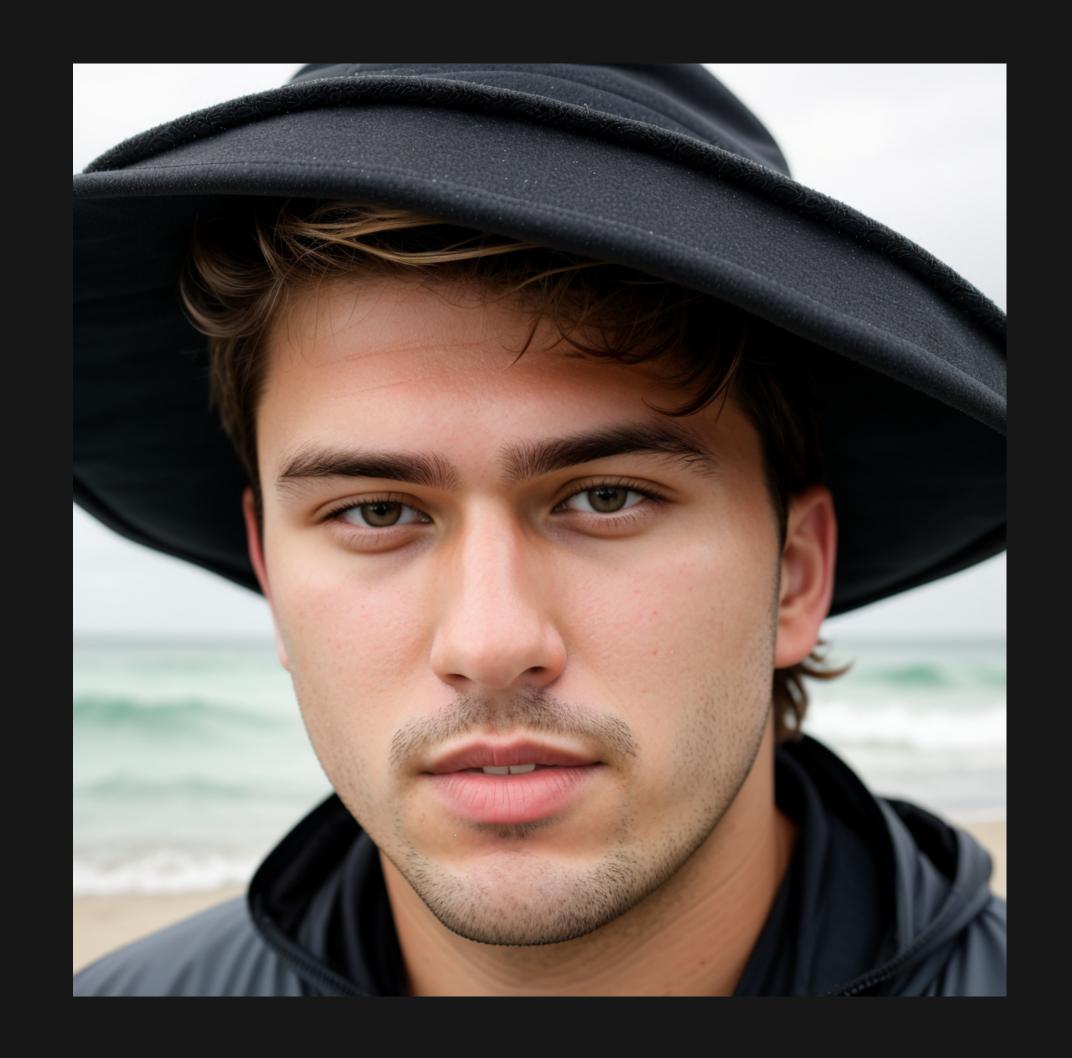
> ...



5. Resolution

How sharp and detailed is the image? Do we want a blurry bokeh background? Etc.

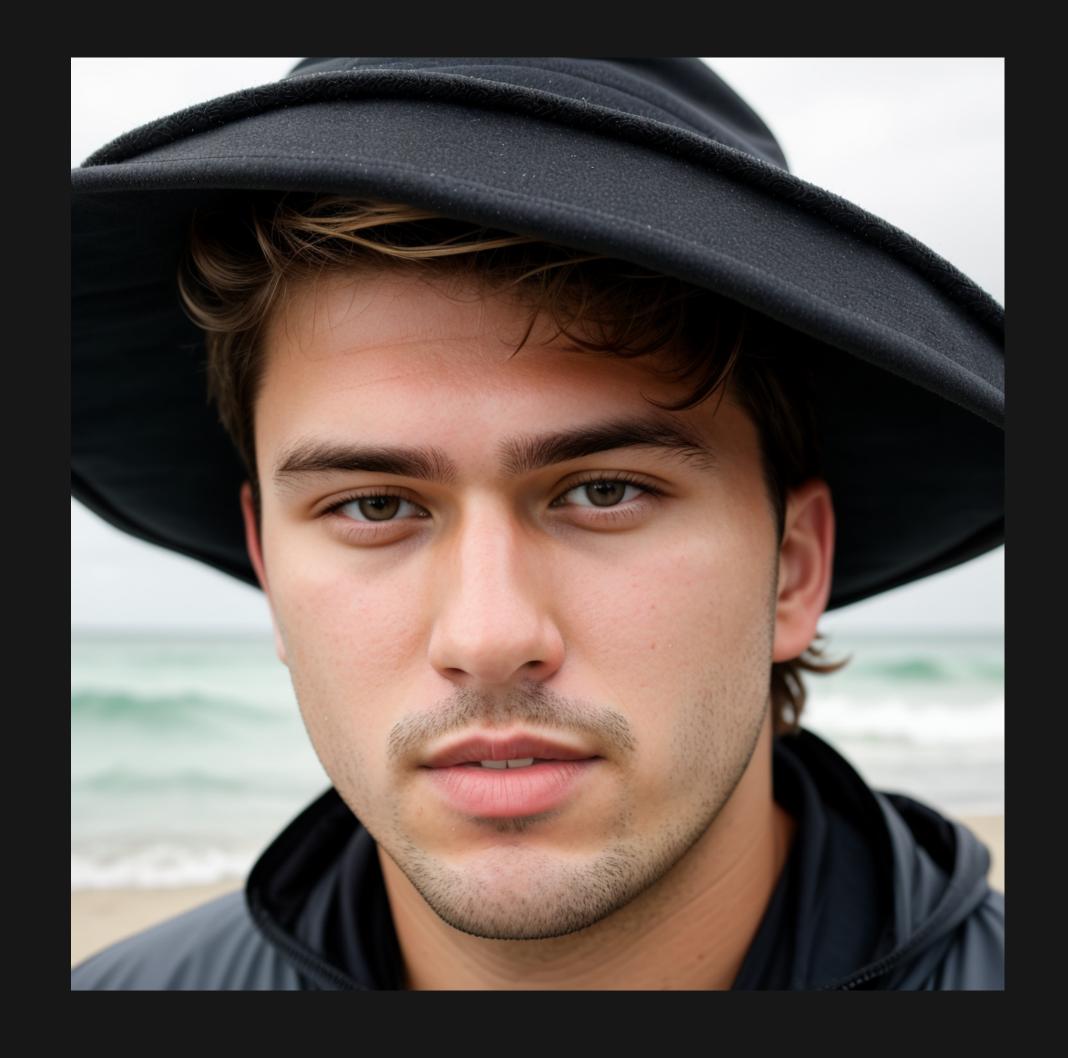
8k UHD, f 1.8, film grain, High detailed skin, skin pores,



6. Additional Details

"Sweeteners" added to modify an image.

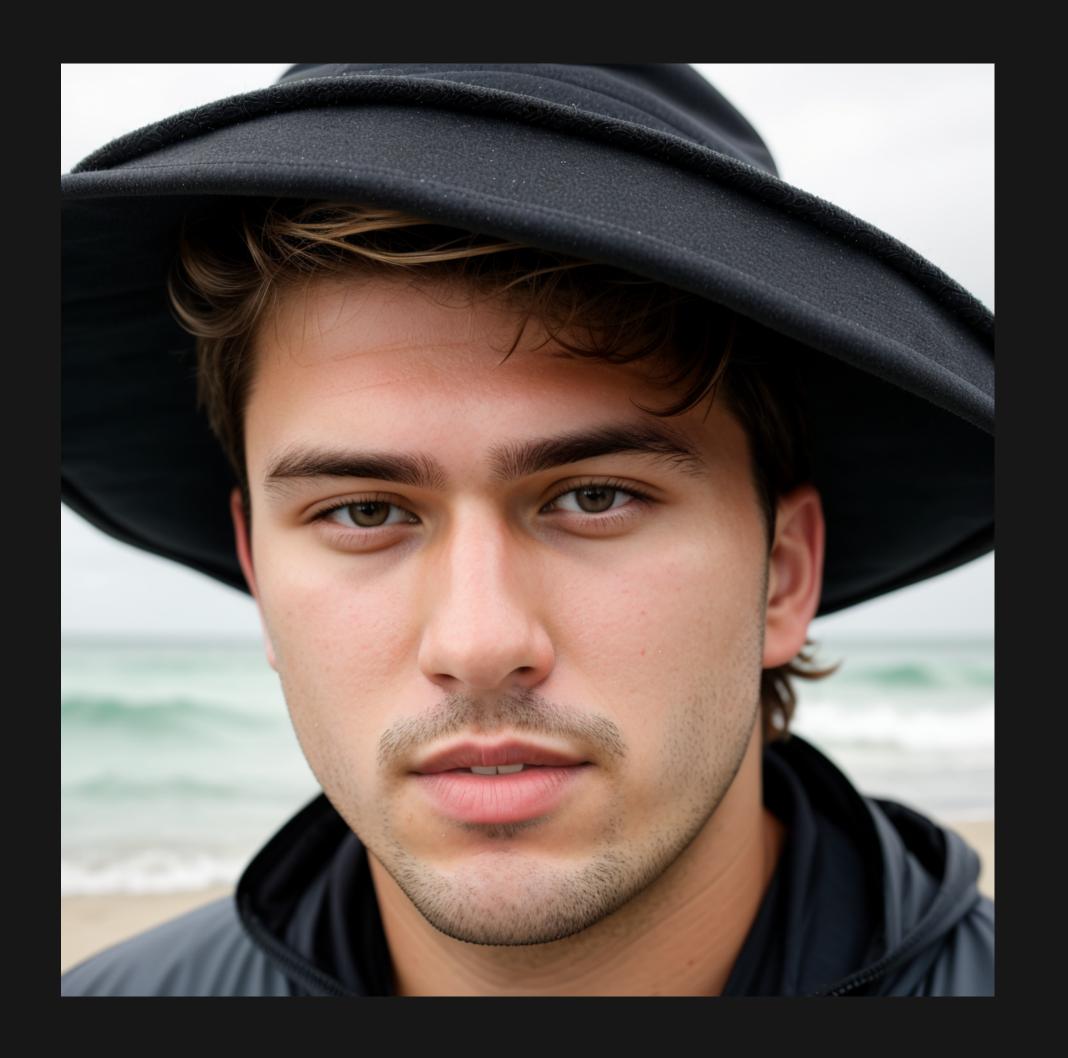
face, skin pores, coastline, overcast weather, wind, waves



7. Color

Vivid, colourful, black & white, sepia, iridescent gold, etc.

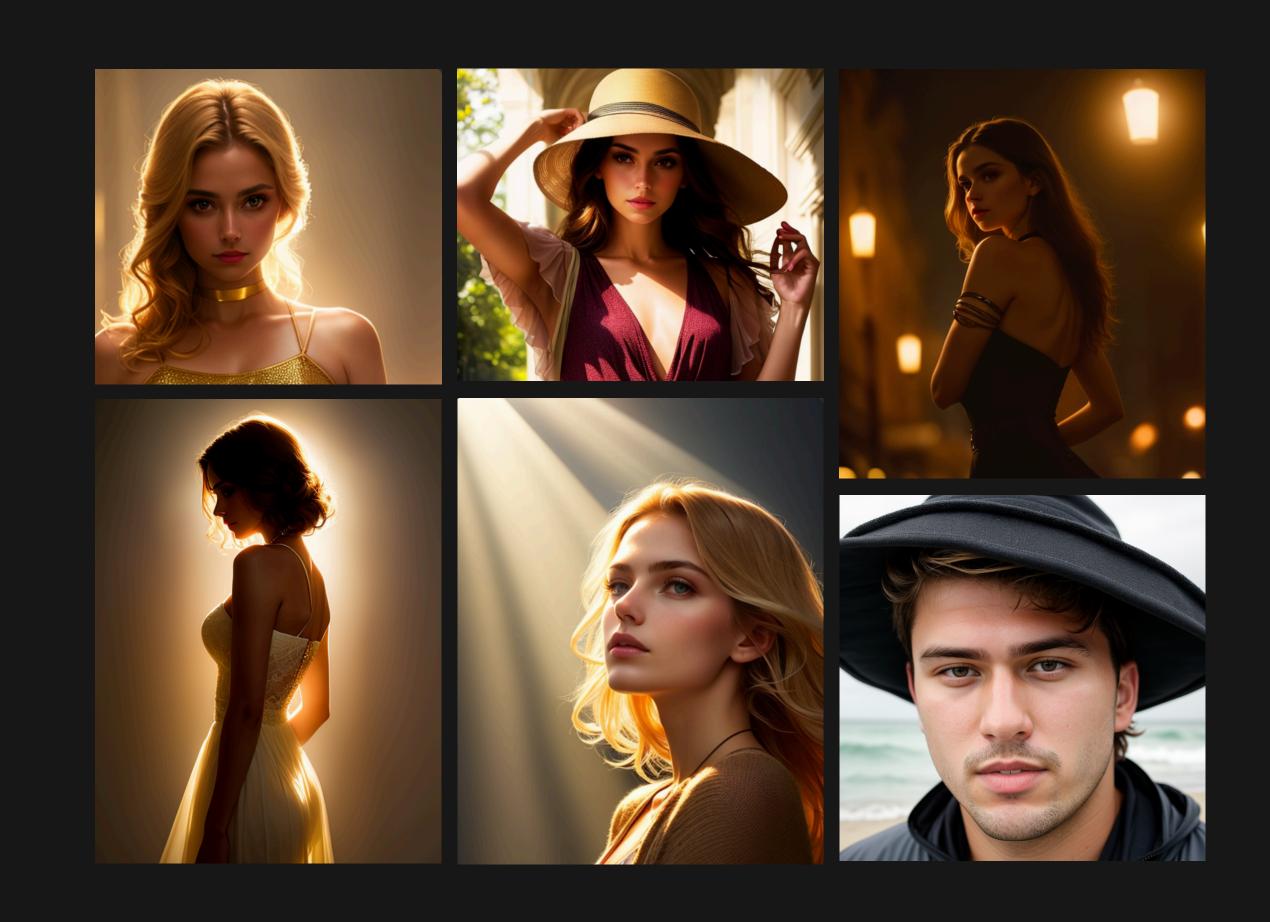
(...)



8. Lighting

Volumetric lighting, rim light, sunlight, backlight, dimly lit, crepuscular rays, soft lighting

> soft lighting (...)

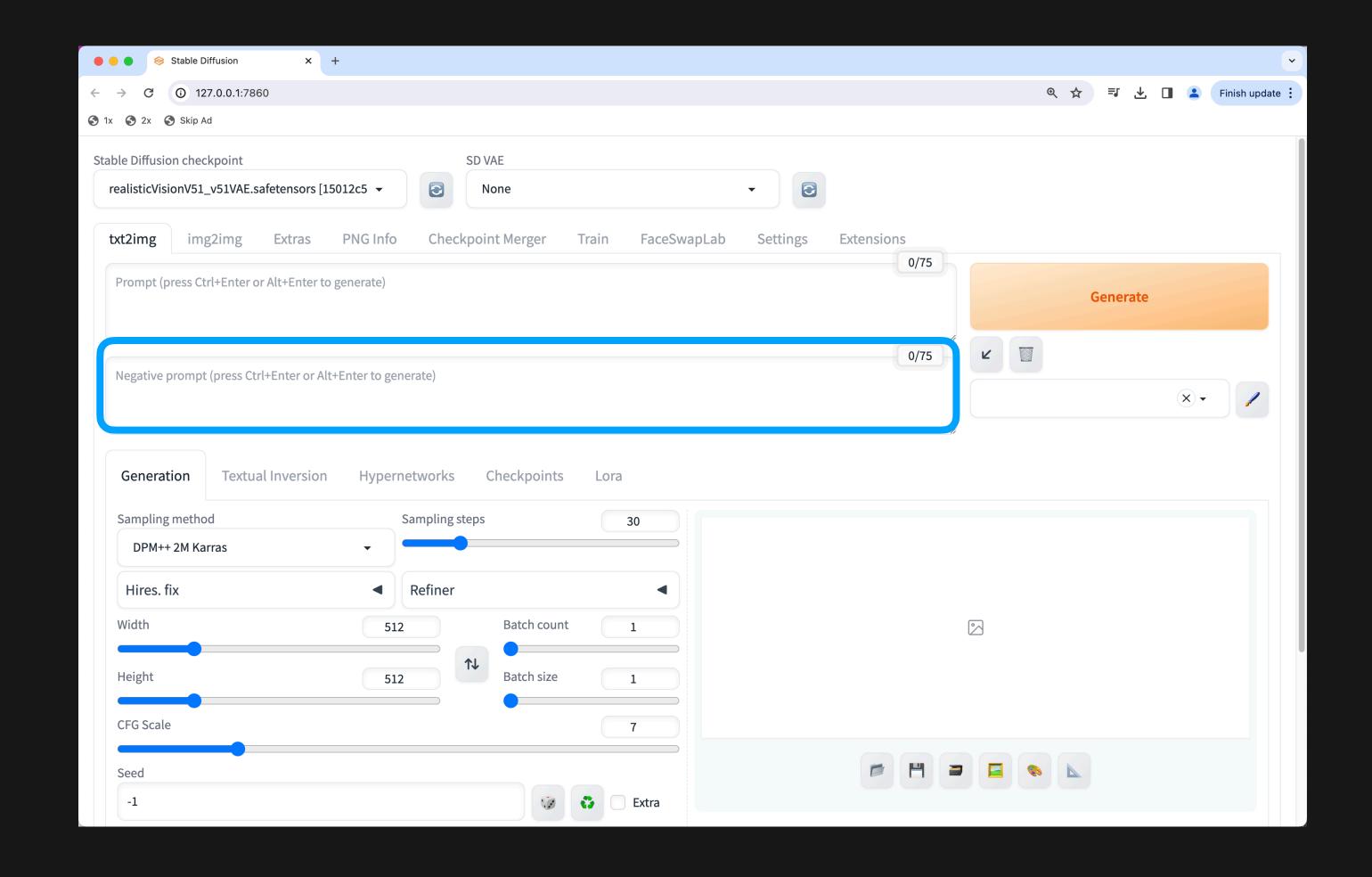


NEGATIVE PROMPT

Here we put what we don't want to see (objects, styles etc.)

Instead of "without a beard" We write "beard" in the negative prompt box.

> ugly, deformed, malformed limbs, squint, etc. TIP: hands



Midjourney uses "--no" at the end of a prompt example: man at coast --no beard, facial hair

KEYWORD WEIGHT

Words at the beginning and at the end of the prompt have more "weight" than the centre.

Another powerful way of giving certain words more/less importance is the syntax:

(Keyword: factor)

I.e. (beard: 0.5) or (beard: 1.5)

Midjourney uses "::number" as prompt weights example: man::2 coast::1 beard::-0.5



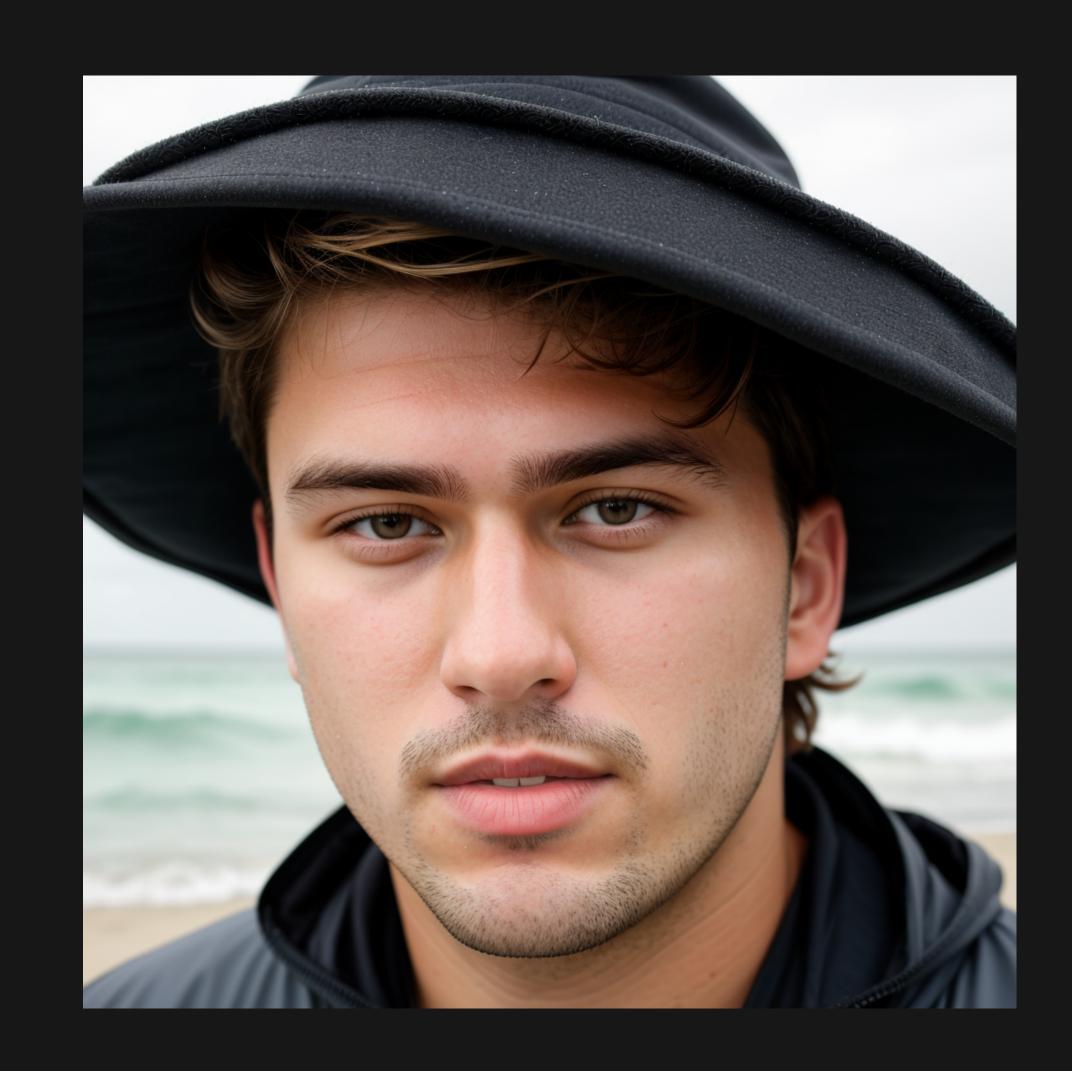




Our final prompt:

RAW photo close-up, 25 y.o. caucasian man in black clothes, face, high detailed skin, skin pores, coastline, overcast weather, wind, waves, 8k uhd, dslr, f 1.8, soft lighting, high quality, film grain, Fujifilm XT3

Negative prompt: beard, moustache





TEXT TO IMAGE IMAGE TO IMAGE

Instead of starting the generation process with random noise we can start it with our custom noise.

You don't need to be an illustrator to create a better starting point than a random starting point. Which is why img2img is way superior.

