

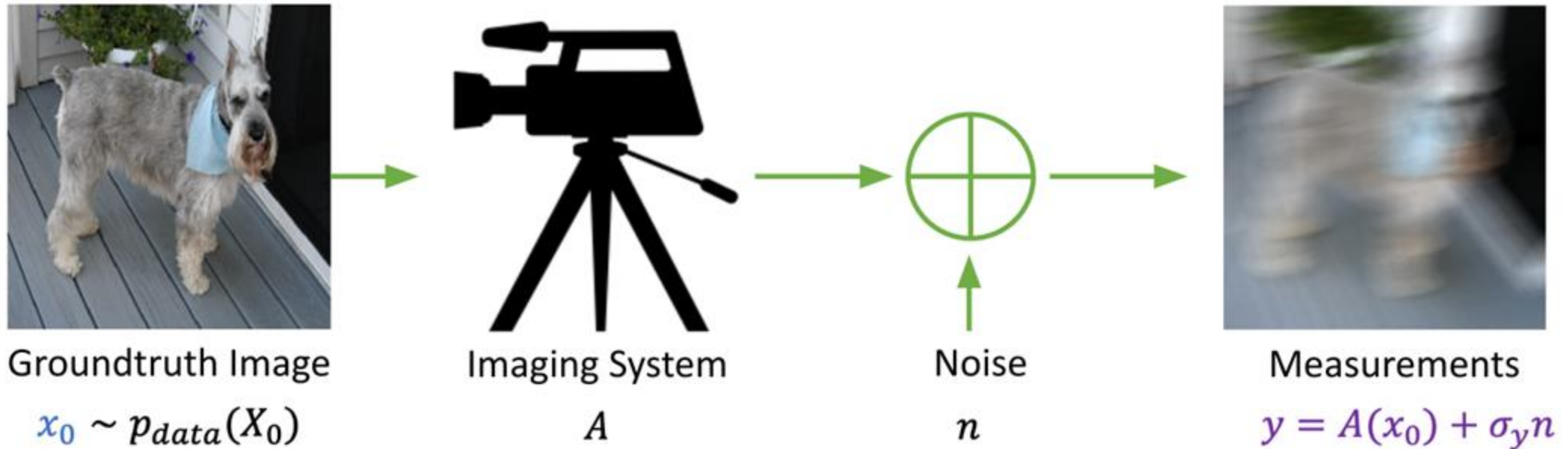
RB-Modulation: Training-Free Stylization using Reference-Based Modulation

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Based on joint work with: Yujia Chen, Nataniel Ruiz, Abhishek Kumar, Wen-Sheng Chu, Constantine Caramanis, and Sanjay Shakkottai

The University of Texas at Austin, Google Research, Google DeepMind

Inverse Problem

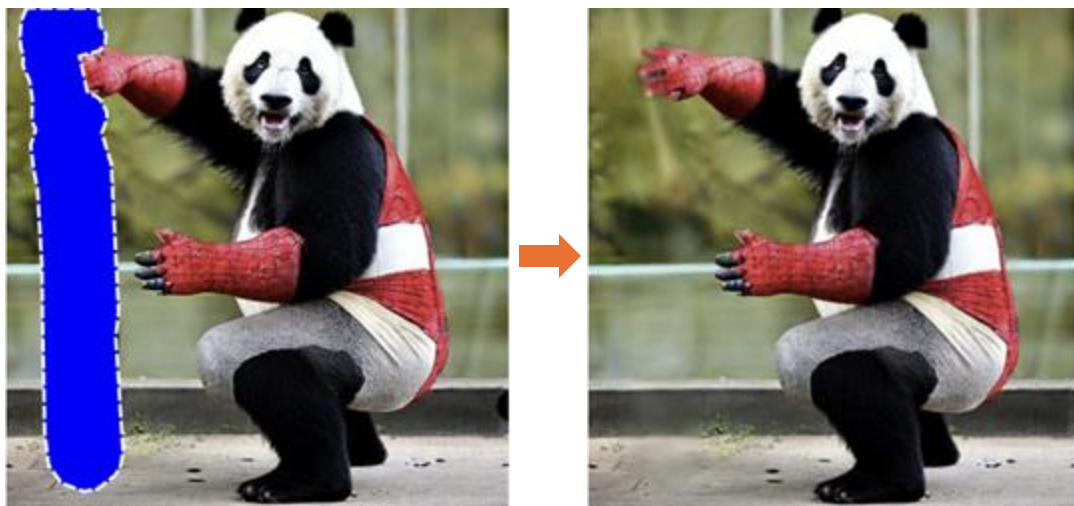


Problem: Reconstruct ground truth image x_0 from noisy measurements y

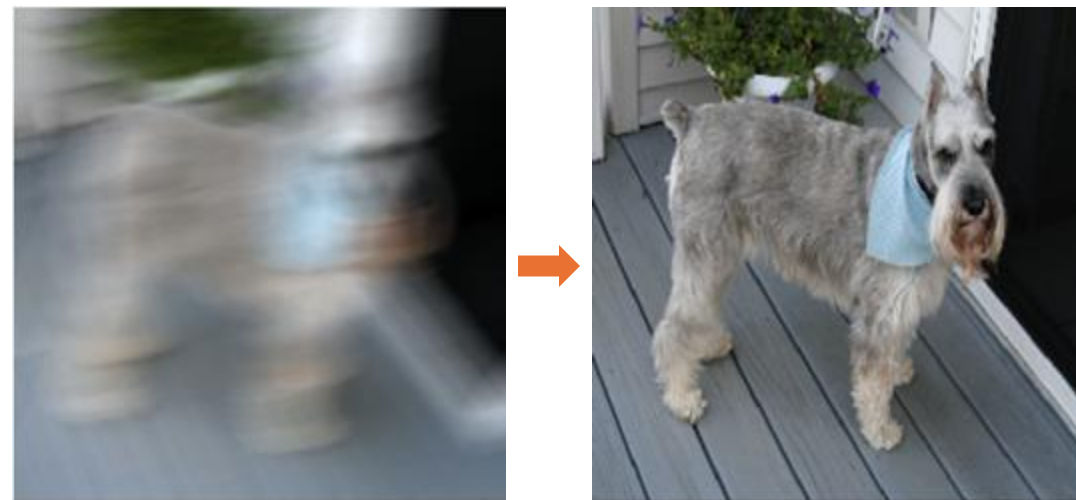
Challenge: Problem is **ill-posed**, that is infinitely many solutions x_0 exist

Approach: Use **prior** knowledge $p(x_0)$ of how the image should look like

General Inverse Problems



Free-form inpainting



Motion Deblur



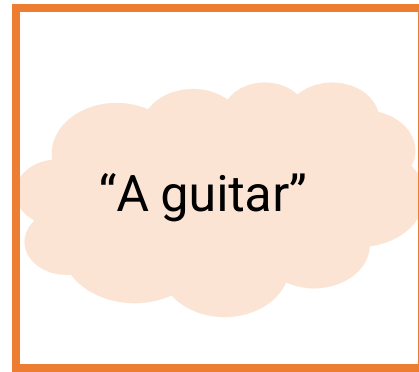
Super-resolution (4X)



Gaussian Deblur

Stylization as Inverse Problem

Text-to-image generation

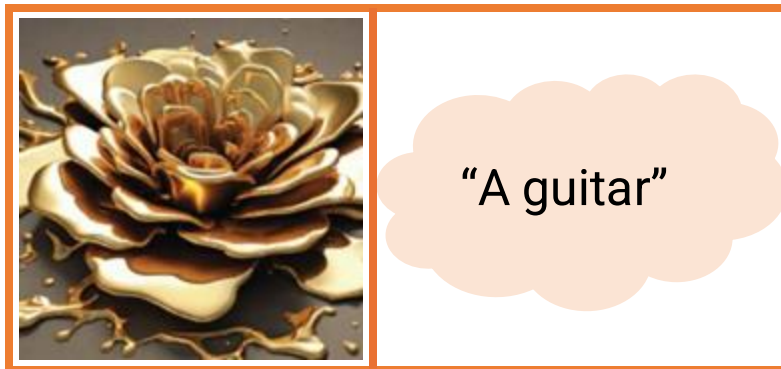


Input



Output

Personalized text-to-image generation: stylization



style

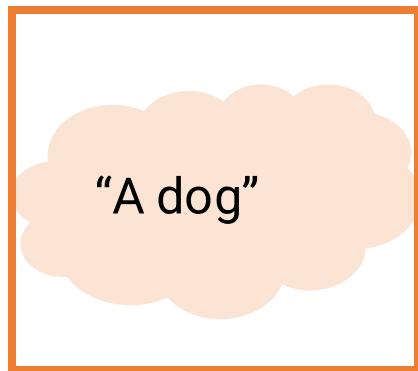
text



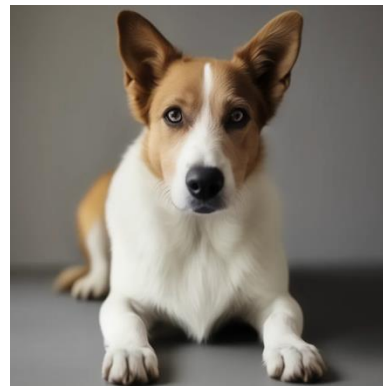
Output

Content-Style Composition as Inverse Problem

Text-to-image generation

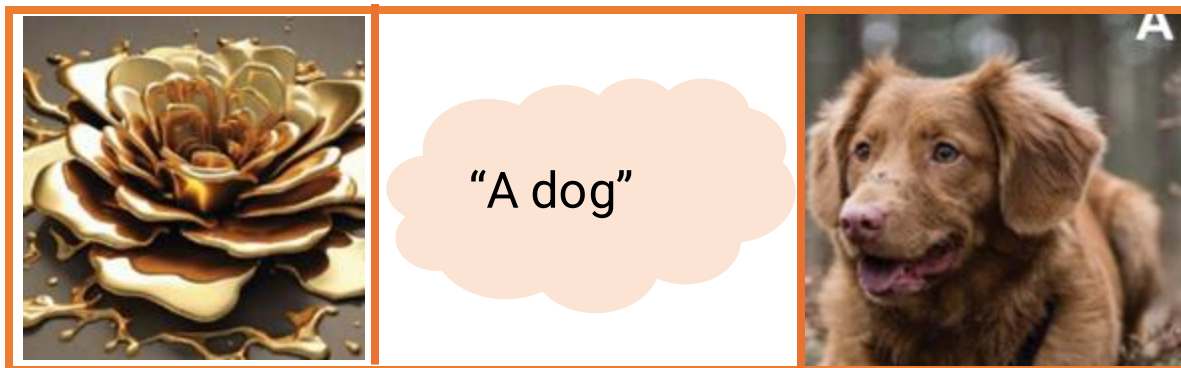


Input

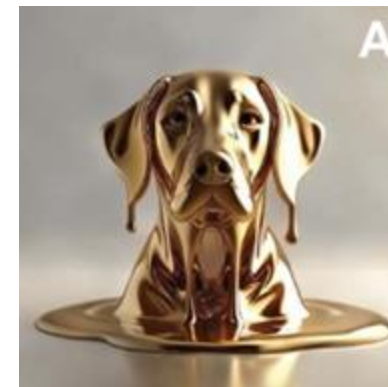


Output

Personalized text-to-image generation: content-style composition



Input



Output

Training-Based Approaches

DreamBooth

- Fully fine-tune the weights of the pre-trained model
- Requires ~4 images per reference subject or style
- Expensive for large-scale text-to-image models
- Catastrophic forgetting due to modified pre-trained weights

Ruiz, Nataniel, et al. "Dreambooth: Fine tuning text-to-image diffusion models for subject-driven generation." *Proceedings of the IEEE/CVF conference on computer vision and pattern recognition*. 2023.

IP-Adapters

- Train newly added cross/self-attention layers
- Requires ~4 images per reference subject or style
- Less expensive for large-scale text-to-image models
- Retains original contents via pre-trained weights

Ye, Hu, et al. "IP-adapter: Text compatible image prompt adapter for text-to-image diffusion models." *arXiv preprint arXiv:2308.06721* (2023).

LoRA

- Train additive residual weights of pre-trained model
- Requires ~4 images per reference subject or style
- Cost effective for large-scale text-to-image models
- Retains original contents via pre-trained weights

Hu, Edward J., et al. "Lora: Low-rank adaptation of large language models." *arXiv preprint arXiv:2106.09685* (2021).

Training-Free Approaches

StyleAligned

- Requires a **text-prompt** describing reference style image
- **Merges** keys and values after instance normalization
- Uses **DDIM inversion** to extract style features from real image
- **Leaks content** from reference style image

Hertz, Amir, et al. "Style aligned image generation via shared attention." *Proceedings of the IEEE/CVF Conference on Computer Vision and Pattern Recognition*. 2024.

Swapping Self-Attention

- Extracts Keys and Values from reference **style image**
- **Swaps** original Keys and Values with those of reference
- Uses **DDIM inversion** to extract style features from real image
- **Leaks content** from reference style image

Jeong, Jaeseok, et al. "Visual Style Prompting with Swapping Self-Attention." arXiv preprint arXiv:2402.12974 (2024).

InstantStyle

- Injects style features through a **specific layer** of an **IP-Adapter**
- Avoids **DDIM inversion** and **memory intensive** reverse SDEs
- Identifying a layer is a complex task and **may not generalize**
- **Limited diversity** due to ControlNet and **leaks content**

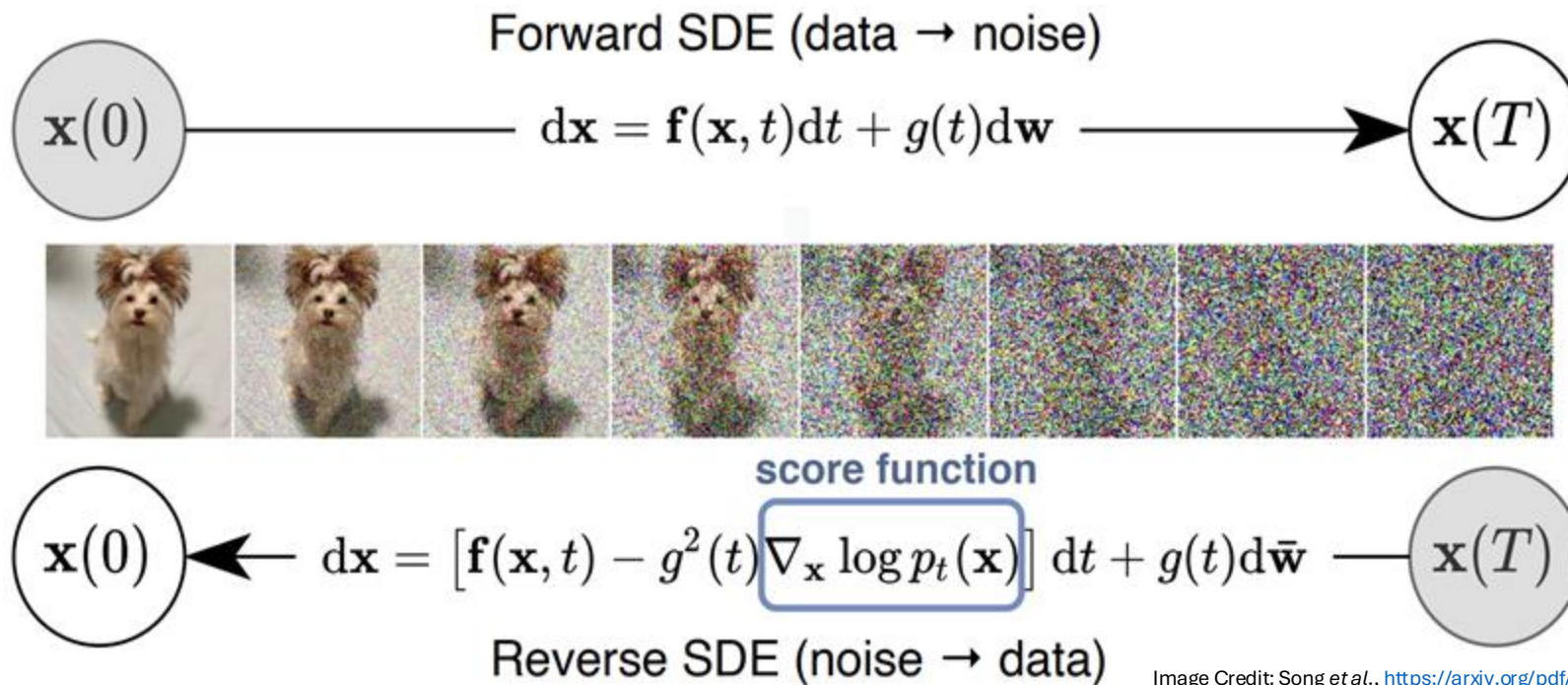
Wang, Haofan, et al. "Instantstyle: Free lunch towards style-preserving in text-to-image generation." arXiv preprint arXiv:2404.02733 (2024).

Our Approach: Modulation of Reverse Diffusion Processes

- RB-Modulation has two key elements
 - **Stochastic Optimal Controller** (SOC) and **Attention Feature Aggregation** (AFA)
- **SOC**: An optimal control formulation-based sampler, implemented as a **test-time optimization algorithm**
- **AFA**: Personalize the score and **disentangle content-style from the reference** images through an alternate cross-attention processor



Background: Diffusion Models



- **Goal:** Design a Markov process-based sampler (a transition kernel) such that stationary distribution samples images
- **Approach:** Learn annealed score that is affine in the conditional expectation of $X(0)$ (clean image) given $X(t)$ (noisy image) by **Tweedie's Formula**

Background: Inverse Problems as Posterior Sampling

Problem: Sample from $p_0(x_0|y)$ instead of $p(x_0)$

$$dX_t = (-X_t - 2 \nabla \log p_t(X_t|y)) dt + \sqrt{2} d\bar{W}_t, t = T, \dots, 0$$

↓
Unknown

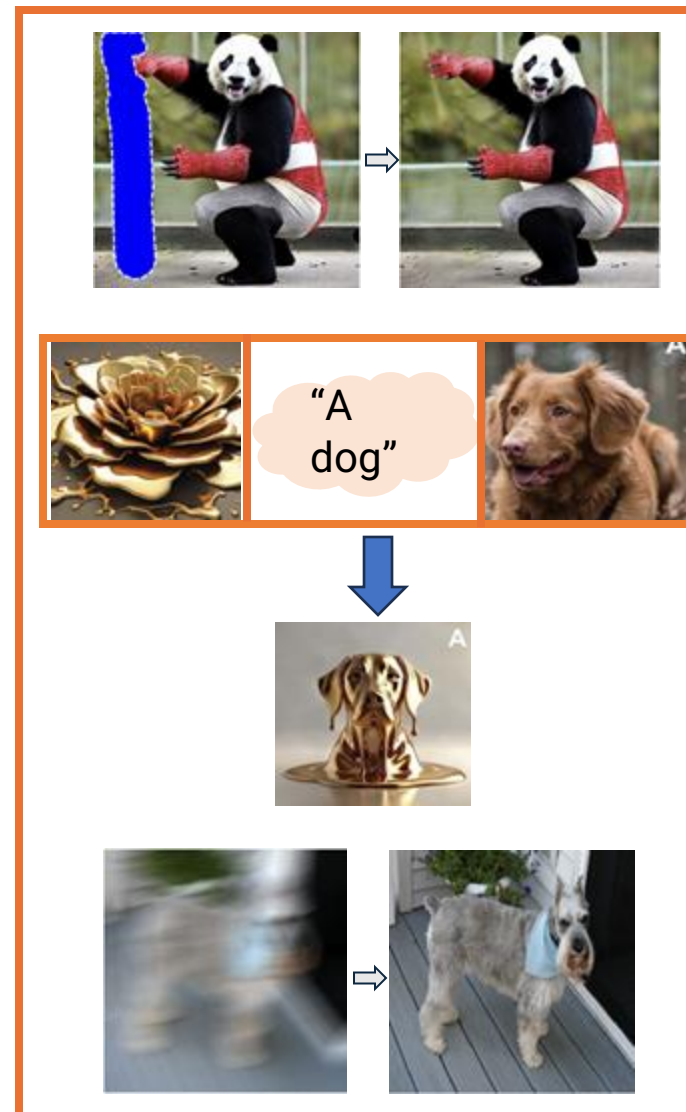
Bayes rule:

$$\log p_t(x_t|y) = \log p_t(y|x_t) + \log p_t(x_t) - \log p_t(y)$$

$$dX_t = (-X_t - 2 \nabla \log p_t(y|X_t) - 2 \nabla \log p_t(X_t)) dt + \sqrt{2} d\bar{W}_t$$

↓ ↓
Unknown Known: $\nabla \log p_t(X_t) \approx s_\theta(X_t, t)$

How well can we approximate $\nabla \log p_t(y|x_t)$?



Our Approach: Posterior Sampling using RB-Modulation

Goal: Interpret posterior sampling as a stochastic control problem

Posterior Sampling: Sample $p_0(\cdot | y)$ using conditional reverse SDE

$$dX_t = (-X_t - 2\nabla \log p_t(y|X_t) - 2\nabla \log p_t(X_t)) dt + \sqrt{2}dW_t, \quad t = T, \dots, 0$$

- Prior approaches^[1,2,3] rely on first- or second-order Taylor's approximation
- We replace $\nabla \log p_t(y|X_t)$ with a controller $u(X_t, t, y)$ and solve a stochastic optimal control problem

$$\min_{u \in U} E \left[\int_T^0 [\|u(X_t^u, t, y)\|^2 + f(X_t^u, t)] dt + g(X_0^u, y) \right]$$

$$\text{where } dX_t^u = (-X_t^u + u(X_t^u, t, y) - 2\nabla \log p_t(X_t^u))dt + \sigma(t)dW_t, \quad X_T^u \sim p_T$$

[1] Chung, Hyungjen et. al. "Diffusino Posterior Sampling for Noisy Inverse Problems", Internation Conference on Learning Representations (2023).

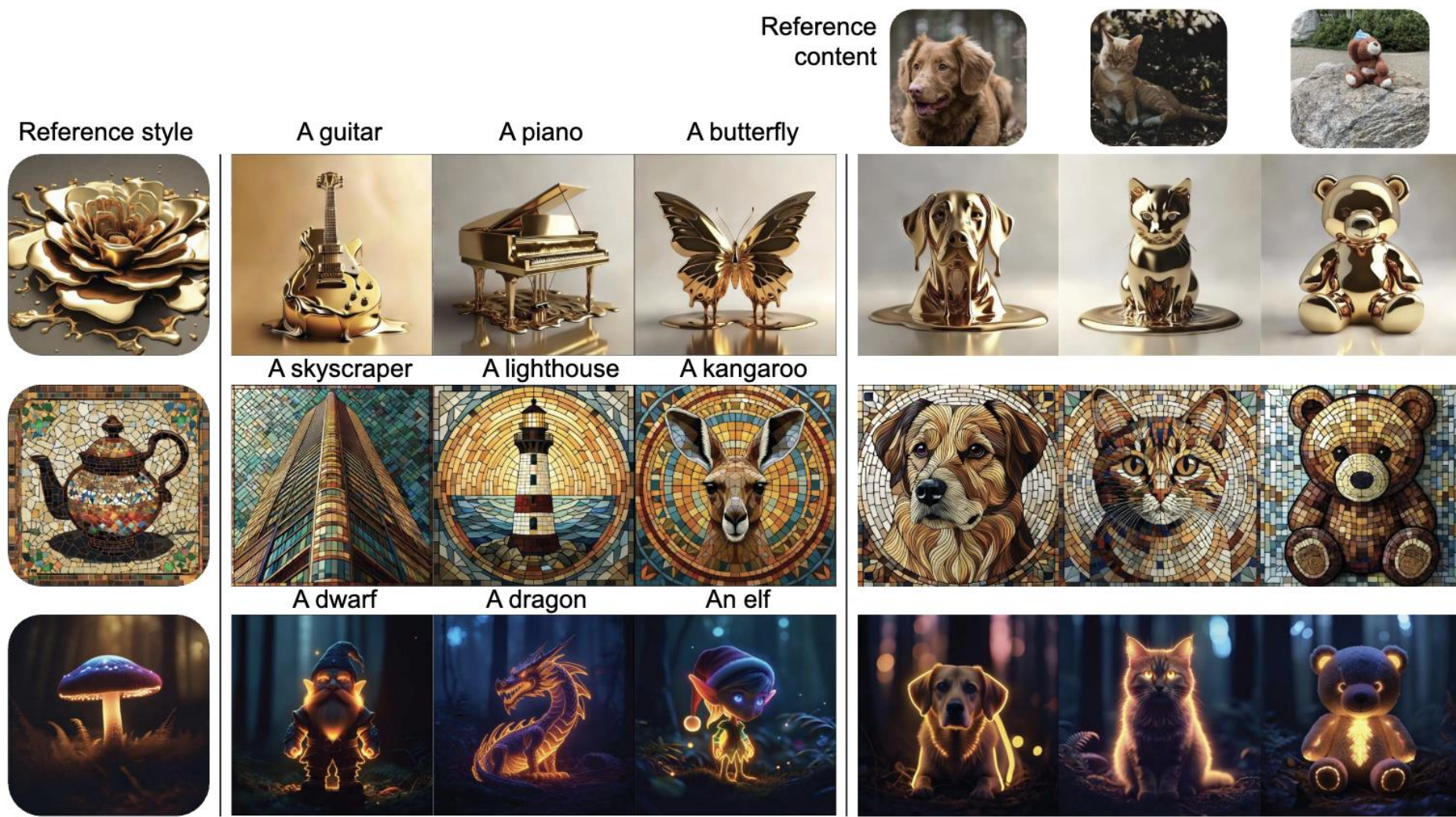
[2] Rout, Litu, et. al. "Solving linear inverse problems provably via posterior sampling with latent diffusion models." Advances in Neural Information Processing Systems 36 (2024).

[3] Rout, Litu, et. al. "Beyond first-order tweedie: Solving inverse problems using latent diffusion." *Proceedings of the IEEE/CVF Conference on Computer Vision and Pattern Recognition*. 2024.

RB-Modulation: Training vs Test-time Optimization

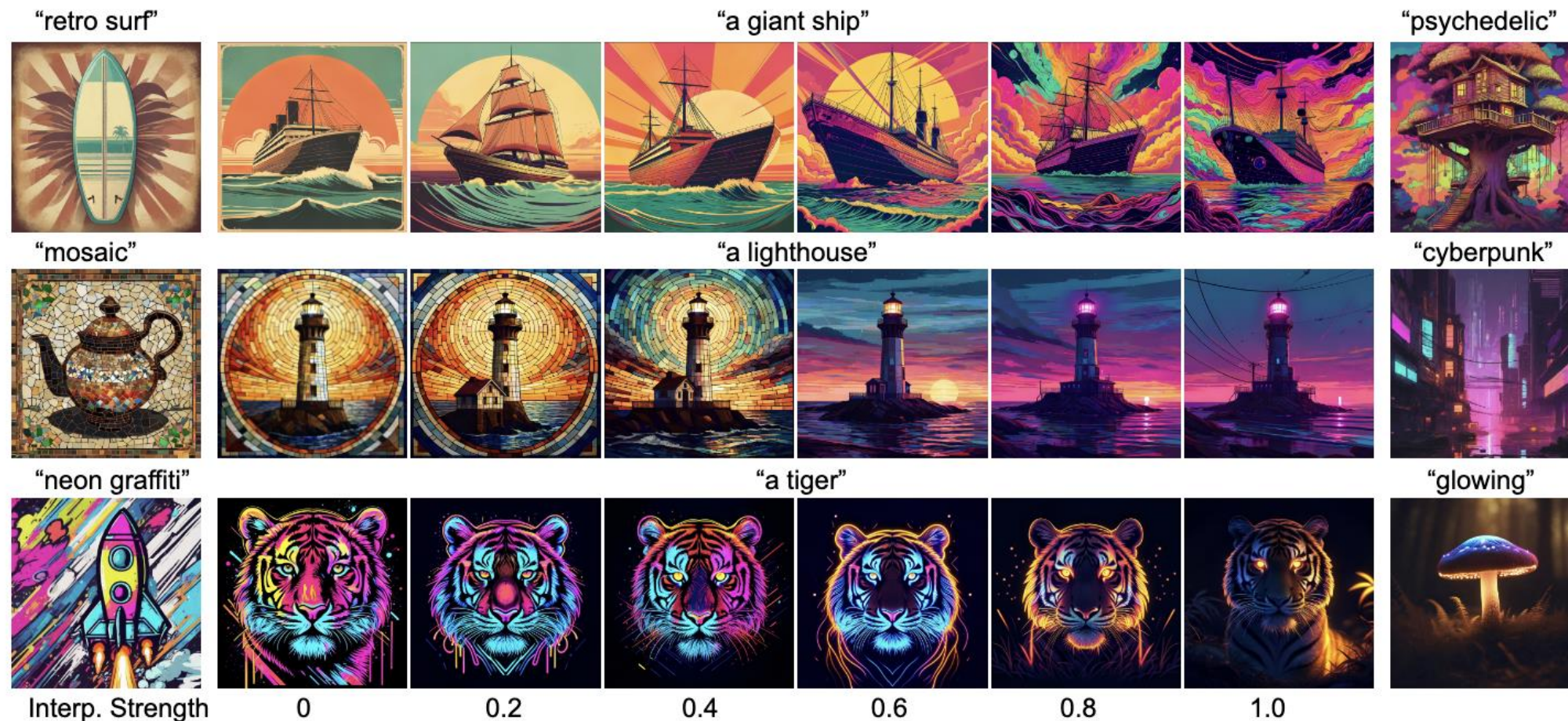
- Training-time optimization (DreamBooth, LoRA, IP-Adapter)
 - Approximately **10s of samples** per conditioning (style/content)
 - Single sample leads to **catastrophic** forgetting
 - Gradient computed with respect to **weights** of base model
 - LoRA finetuning takes **~20** min per style (40 min for content-style)
 - Full finetuning takes **hours**
- Test-time optimization (DPS, **PSLD**, P2L, **STSL**)
 - **Single** sample suffices (no catastrophic forgetting)
 - Gradient computed with respect to **input** to base model
 - Takes **~10** min for PSLD (1B), **~20** min for P2L(1B) (longer for Flux-12B)
- **Proximal** test-time optimization (**RB-Modulation**)
 - Takes **40 sec** using StableCascade (4B)

Experiments: Training-free Personalization



RB-Modulation as a plug-and-play solution for (a) stylization (b) content-style composition

Novel Style Synthesis: Interpolating Reference Styles



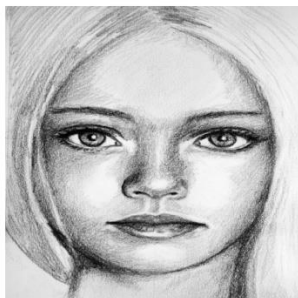
Training based methods cannot interpolate novel styles
due to lack of prior examples

Stylization: Hand Drawn Reference Images

“plastic crayon”



“pencil sketch”



“comm. paint”



Reference Style “house on a mountain”

“racing car”

“futuristic robot”

“tiger”

“lion”

Project page: <https://rb-modulation.github.io/>

RB-Modulation: Training-Free Stylization using Reference-Based Modulation

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ICLR 2025 (Oral: 1.8% acceptance ratio)

[\[Paper\]](#)

[\[OpenReview\]](#)

[\[ArXiv\]](#)

[\[Code\]](#)

[\[Demo\]](#)

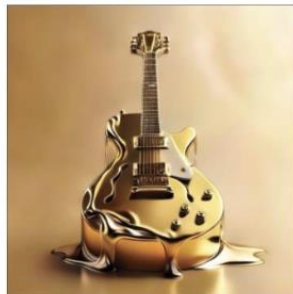
Reference
content



Reference style



A guitar



A piano



A butterfly



A skyscraper

A lighthouse

A kangaroo

